PHYSICS

A Textbook for Higher Secondary Schools

Classes XI-XII

Part II

(Volume 2)

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National Council of Educational Research and Training

UNIT 16

Solids

16.1 Structure of Solids

Man has used solids from a very ancient period when he was dwelling in caves but it is only in this century that he has gained a satisfactory understanding of the way constituent atoms arrange themselves in a solid. You have come across commonly occuring crystals of sugar, rocksalt, mica, quartz and some gems. The crystals have symmetrical shapes and flat faces which meet at characteristic angles. Most of the solids, which do not appear to be crystalline to the naked eye, are actually found to be aggregates of small crystals when examined under a microscope. Questions like why diamond is hard, what makes copper a good conductor of electricity and iron a magnet, naturally occur to us. The first step towards answering such questions and understanding the nature and behaviour of solids is the knowledge of the structure of solids i e the knowledge about the arrangement of atoms in a solid.

The diffraction of X-rays by crystals demonstrated directly that a crystal is a periodic arrangement of atoms or molecules. We had already noted while discussing diffraction of waves that a periodic array of atoms in three dimensions would readily diffract X-rays whose wavelength is of the same order as the interatomic distance. The X-ray diffraction has been successfully used to probe the structure of variety of crystals ranging from simple materials like sodium chloride and copper to some very complex materials like proteins.

To describe the structure of a crystal we do not have to give positions of all the atoms in the crystal. The regularity and the orderliness which is there in the atomic arrangement makes this problem quite simple. The arrangement of atoms in a crystal can be described by choosing a small group of atoms and imagining them to be located in a cell. Repeating this unit cell along three definite directions in space provides the atomic positions in a crystal. We illustrate this in Fig. 16.1 by a two dimensional example for the sake of simplicity even though strictly speaking there does not exist, in the real world, a two dimensional crystal. Here the repeating unit is chosen to consist of two atoms represented by white and black circles. This is only a schematic representation and is not intended to show the shape and size of atoms faithfully. The choice of the unit cell is quite arbitrary. Two possible choices are shown in Fig. 16.1. The knowledge of the location of atoms in the unit cell contains complete information regarding the structure of the crystal. This introduces an extreme simplification in the study of crystals because instead of considering the whole crystal, one has to consider only the unit cell.

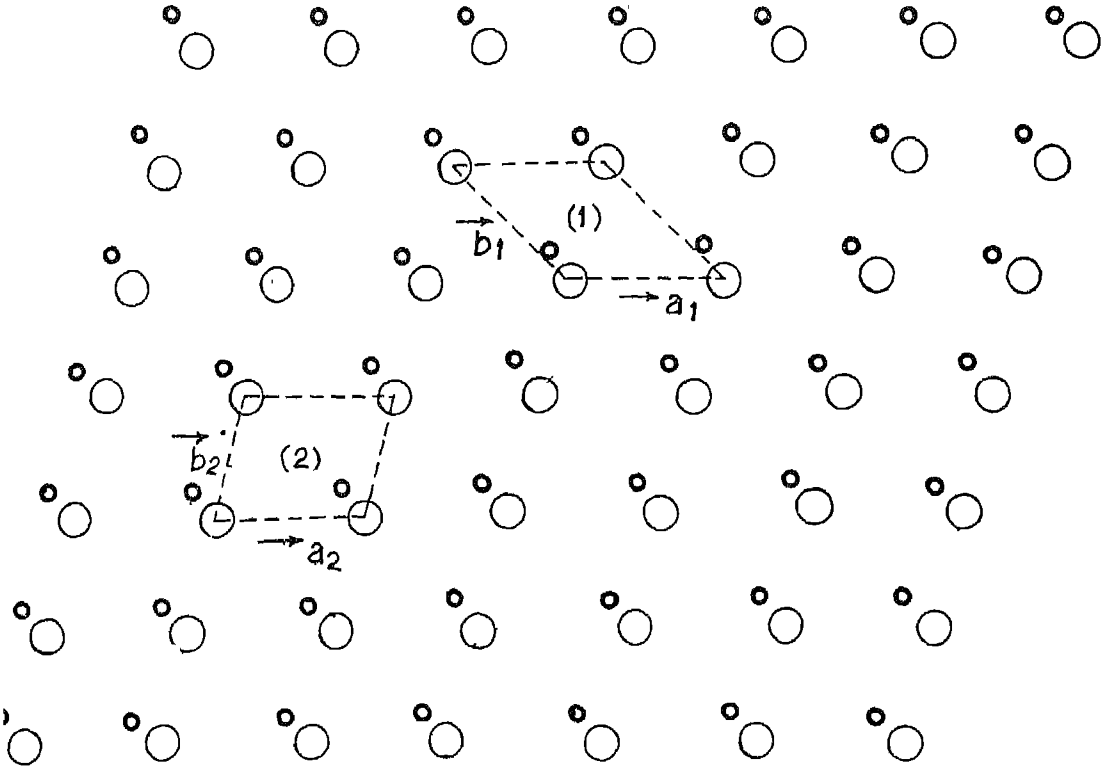

Fig. 16.1. The ordered periodic arrangement in crystals The choices (1) and (2) of the unit cell are shown

16 2 Crystal Lattice

The study of the geometry arrangement of atoms in a crystal becomes simple if we replace the crystal by a lattice. A lattice is a purely geometrical concept whereas crystal is a real thing. We explain the concept or the lattice with the help of the two dimensional structure as shown in Fig. 16 1. There we have a group of two atoms repeating themselves according to a pattern. If we associate a point with this group and replace all the groups of atoms by points placed at corresponding positions, we get a collection of points (Fig 16 2). The array of these points is known as the lattice. A lattice is thus a regular arrangement of points in space The crystal structure is obtained by attaching the group of atoms to lattice points. The lattice points are given by the tips of lattice vectors :

$$\vec{l} = n_1 \vec{a} + n_2 \vec{b} \qquad \ldots(16.1)$$

with some arbitrary origin O. $\vec{a}$ and $\vec{b}$ are *translation* vectors, and are shown in Fig. 16.2 for the two choices (which are indicated there by the subscripts to $\vec{a}$ and $\vec{b}$). n_1 and n_2 are arbitrary integers.

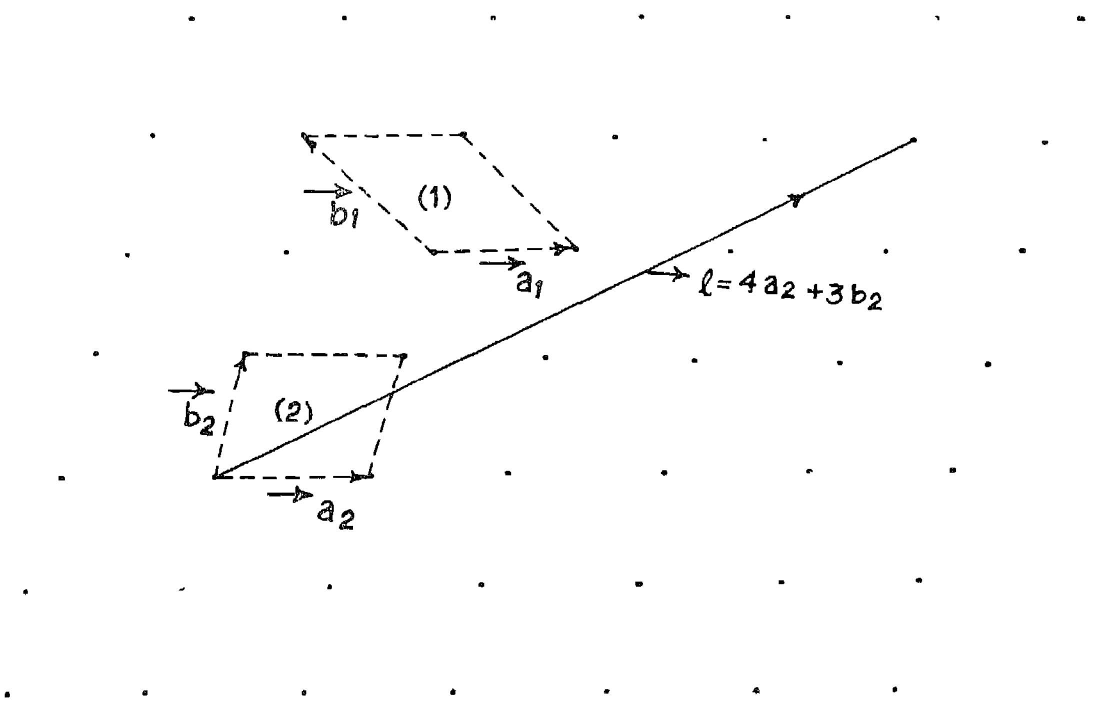

Fig. 16 2. The lattice corresponding to crystal structure shown in figure 16.1.

The Eq 16.1 can be generalized to three dimensions where the lattice points of a three dimensional lattice would be given by

$$\vec{l} = n_1 \vec{a} + n_2 \vec{b} + n_3 \vec{c} \quad ...(16\ 2)$$

16.3 Amorphous Solids

We have seen above that the regular periodicity in the arrangement of atoms is the characteristic feature of *crystalline solids*. There is no such orderliness found in the arrangement of atoms in amorphous solids. Glass is an example of an amorphous solid In fact, glass is used as a synonym for amorphous solid. Like liquids the amorphous solids have a disordered arrangement of constituents (molecules or atoms). The main difference in liquids and amorphous solids is that in liquids molecules are free to move within the volume of the liquid whereas in the amorphous solids the atoms are rigidly fixed.

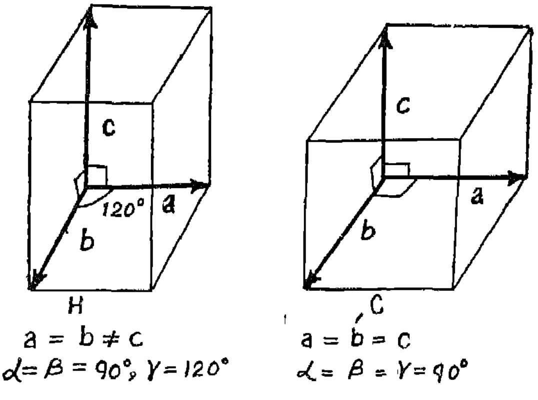

Fig 16.3. Crystal systems : H—hexagonal C—cubic.

The glass does not have a well defined freezing point. We know that when a liquid crystallizes it does so at a definite freezing point at which the heat of solidification (latent heat of freezing) is released. This corresponds to the lowering of the energy of the system due to the ordering of the atoms of the system to the regular arrangement in a crystal. In glasses the freezing into a crystalline state is prevented either due to high viscosity of the liquid or due to the fast rate of cooling. A glass, therefore, retains random disordered arrangement similar to that of a liquid.

16.4 Crystal Systems

We have seen in Fig 16.2 that a variety of unit cells can be chosen for a crystal lattice. All crystal lattices can be grouped into seven crystal systems according to the seven different types of unit cells we can choose for them.

The unit cells for two of these seven systems "the cubic and the hexagonel" are shown in Fig. 16.3. Unit cell is specified in terms of the translation vectors $\vec{a}$, $\vec{b}$ and $\vec{c}$ and the angles α, β, and γ, between them. The notation used for the angles is explained in Fig. 16 4.

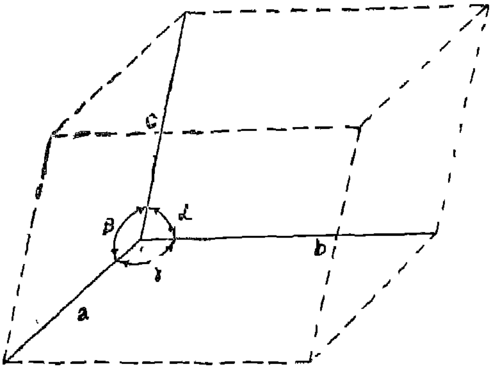

Fig. 16 4 A unit cell defined by vectors $\vec{a}$, $\vec{b}$, $\vec{c}$. The angles α, β and γ are shown, for example α is the angle between $\vec{b}$ and $\vec{c}$.

16.5 Close Packed Structures

Many crystals found in nature, specially those of common metals and inert gases (the inert gases also crystalize at low enough temperatures) can be viewed in terms of packing of spheres, where the atoms in the solid are likened to hard impenetrable spheres The two ways of close packing spheres (atoms) in three dimensions are discussed here. If we try to pack spheres as closely as possible in a plane, there is only one way shown in Fig 16.5, where spheres are sitting at positions marked, A, forming an equilateral-triangular array. Here each sphere in the plane is surrounded by six spheres which just touch it. A second layer of spheres over this layer is placed so that each sphere (B) of this layer settles over a depression (centered at the point indicated by +) formed by three spheres in the first layer as shown in Fig 16.6. A third layer can now be added in two *distinct* ways. In

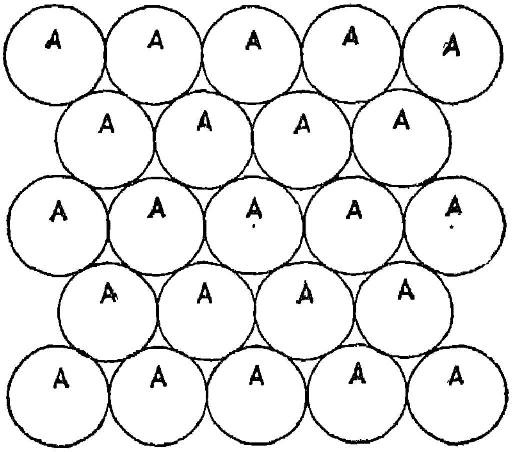

Fig. 16 5. The close packing of spheres in a layer.

one, the spheres in this layer are directly over the corresponding spheres in the first layer and we call this sequence of these three layers ABA. In the second arrangement

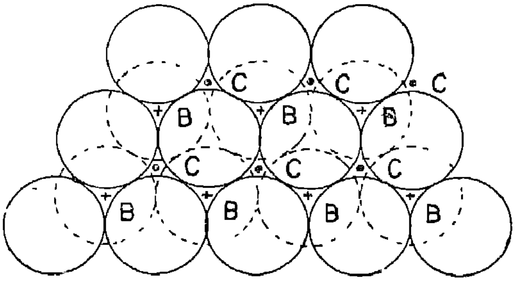

Fig 16.6 The second layer of spheres is shown dashed.

the spheres in the third layer lie directly over the hollows marked C of the first layer. We call this sequence ABC. We note from Fig 16 6 that hollows C are the ones which were left out when we placed spheres in the second layers at hollows marked B The continuation of the above two ways of stacking spheres in space, ABABAB .... and ABCABCABC ...., leads to two types of the close packed structures known as *hexagonal close packed* (hcp) structure and *face centred cubic* (fcc) structure respectively A typical example of hcp structure is magnesium Aluminium, copper, silver, gold, nickel, platinum and many other metals possess fcc structure The position of atoms in a hcp structure is shown in Fig 16 7. The unit cell for the structure has been outlined with heavy lines. A similar figure could be drawn for the fcc structure

The above are the only two ways of arranging atoms in close packed structures. Compared to these, other structures are loosely packed. We see that in close packed structures there are twelve nearest neighbours to each atom, six in its plane, three in the plane above and three in the plane below. This number of nearest neighbours known as the *coordination* number is less for structures which are not close packed.

16.6 Forms of Binding in Solids

The properties of a solid clearly depend on the kind of binding that exists between atoms. The forces which bind atoms together in solids are all electrostatic in origin. The strength and the type of the binding is determined by the distribution of the electrons in the outermost shells in the atoms

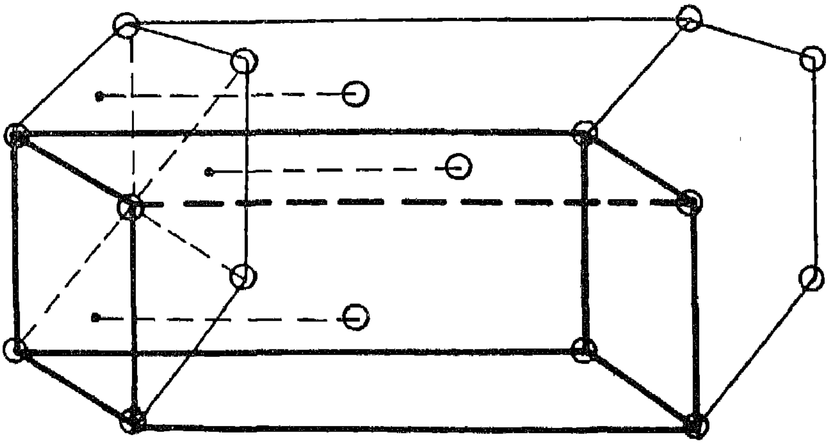

Fig 16.7. Atomic positions in the hcp structure.

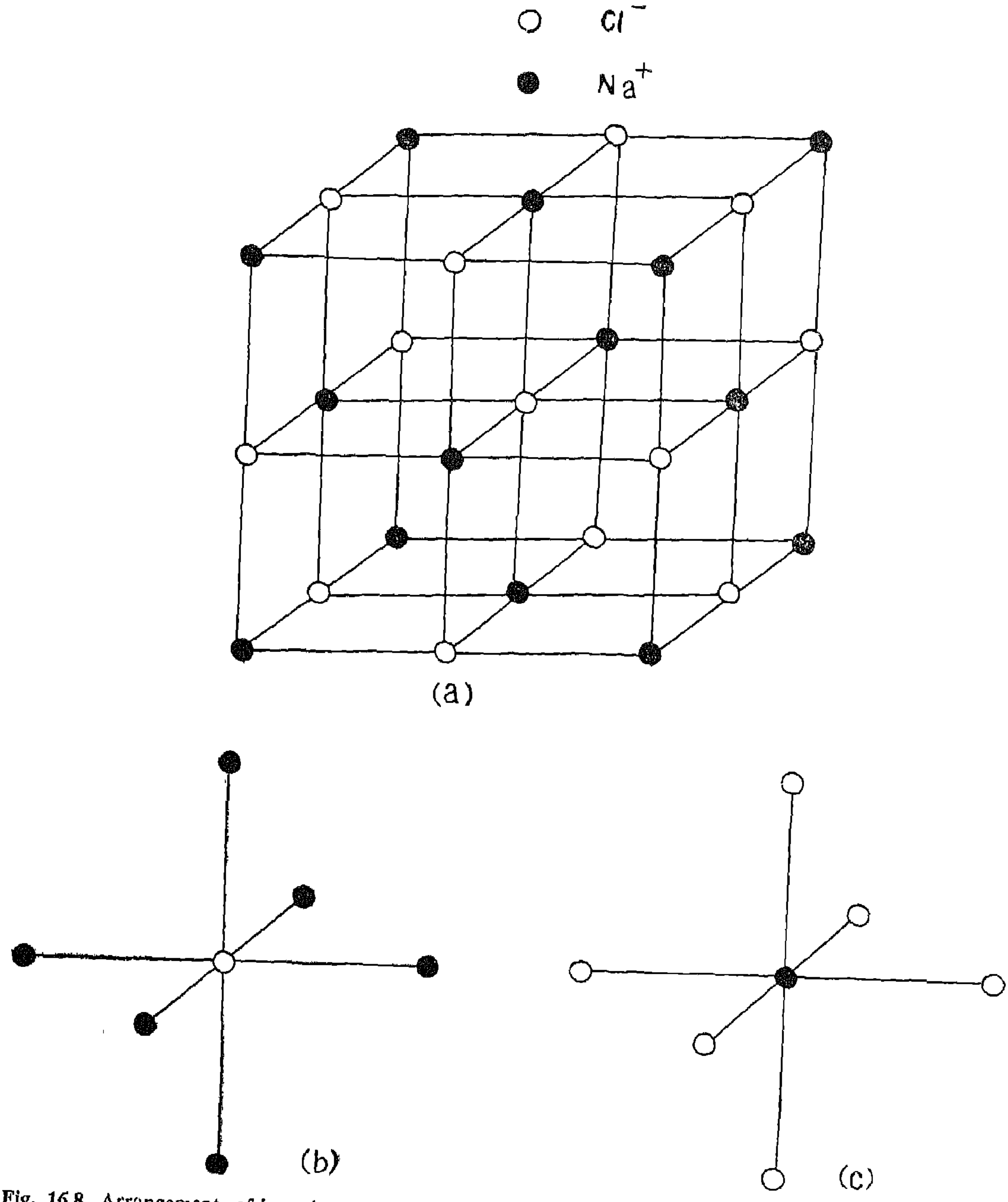

Fig. 16 8 Arrangement of ions in NaCl crystal. The figures (b) and (c) show nearest neighbours of a Cl^- ion and a Fa^+ ion respectively.

involved. The binding in crystals is conveniently divided into four types ionic, covalent, metallic and Van der Waals. We shall discuss each of these types of binding separately.

The Ionic Binding

If we bring together equal number of sodium and chlorine atoms, it is easy for the valence (3s) electron of sodium to be transferred to the 3p shell of chlorine and we have Na$^+$ and Cl$^-$ ions. Both these ions have stable closed shell structures (Na$^+$ = $1s^2\,2s^2\,2p^6$ and Cl$^-$ = $1s^2\,2s^2\,2p^6 3s^2\,3p^6$) similar to those of inert gas atoms In common salt (sodium chloride) the positive ions of sodium (Na$^+$) and the negative ions of chlorine (Cl$^-$) are arranged in the manner shown in Fig 16 8. In this arrangement the nearest neighbours of any ion are ions of opposite charge (Fig. 16.8b and Fig 16.8c). It is for this reason that attractive forces between opposite charges predominate over the repulsive forces between the like charges. If the effective force remained attractive for any separation of ions, the ions would collapse into each other This is prevented due to the repulsive force that comes into play when the separation between ions is decreased.

Many other crystals, like compounds formed from an element of group I of the periodic table and some element of group VII, are held together by this ionic binding and these crystals are known as ionic crystals. The cohesive energy of an ionic crystal is defined as energy required to transform the ionic crystal into a system of isolated ions (Fig. 16.9). The higher the value of the cohesive energy the more stable the crystal would be. Ionic crystals have high cohesive

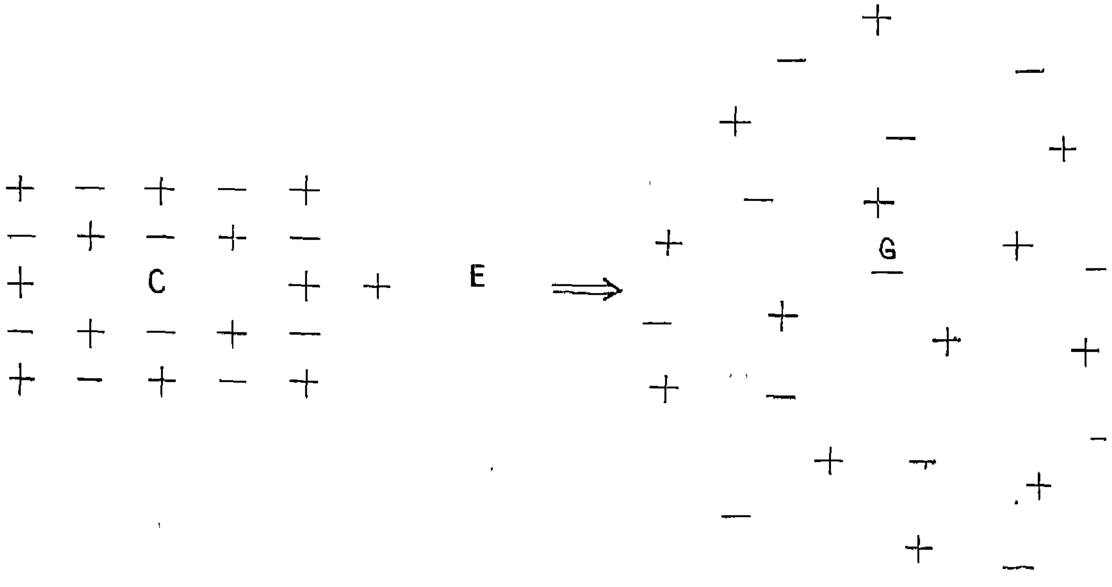

Fig. 16.9. Schemetic illustration of cohesive energy of an ionic crystal. C—crystal, E—cohesive energy, C—gas of isolated ions.

energy, therefore the melting points of ionic crystals are also high. Because all the electrons in these crystals are bound to that ion core, there are no free electrons to carry electric current, such solids are therefore poor conductors of electricity.

Covalent Binding

In this class of crystals the constituent atoms are bonded together by sharing of electrons with neighbours Group IV crystals like diamond, silicon, germanium have covalent bonds. For example, in diamond, a carbon atom is involved in four covalent bonds with its four nearest neighbours which lie at the vertices of a tetrahadron centred at the atom under consideration (Fig. 16.10). This is a loose packed structure with coordination number four The four outer electrons of the carbon atom can join with electrons of adjacent four carbons give a strong directional bond. Each covalent bond is formed by two electrons and each carbon atom sharing the bond contributes one electron to it. The electron charge density for these two electrons participating in the bond is concentrated along the bond in between the two atoms. This is shown schematically in Fig. 16.11. The neighbouring bonds repel each other and the four bonds at tetrahedral angle (109.5°) are directed towards the four vertices of the tetrahedron The directional nature of the covalent bonds does not allow the crystal to assume the close packed structure. Crystals with covalent bonds are very hard. The materials are difficult to deform due to directional nature of the bond.

Many solids are partly ionic and partly covalent. The ionic character progressively increases in the sequence of solids Ge (IV group element)→ GaAs (III—V compound)

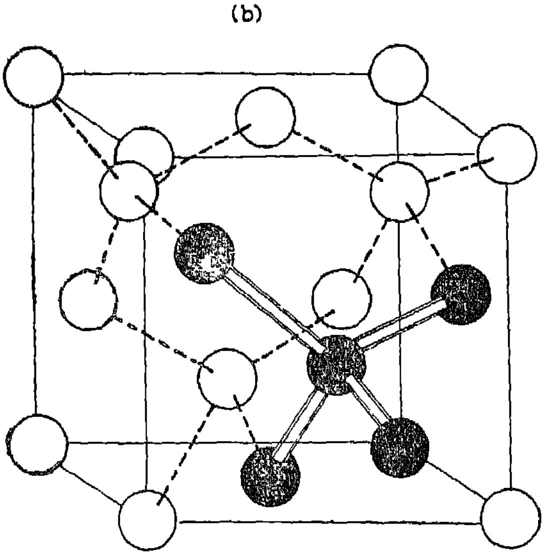

Fig 16.10 The unit cell of the diamond showing tetrahedral bonding in the crystal. One of the tetrahedral units have been marked out by denoting the carbon atom involved in this unit by filled circles and the bonds by double lines.

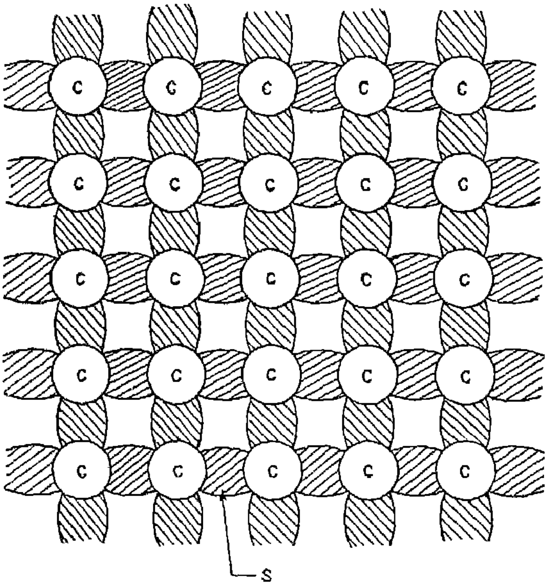

Fig. 16.11. Highly schemetic representation of the electron distribution in covalent bonding in diamond S—shared electron.

→ZnSe (II—VI compound) →CuBr (I—VII compound) For example, in ZnSe all the electrons are not shared, Zn and Se are partly charged due to permanent transfer of some electrons from Zn to Se so that the crystal is partly ionic.

Metallic Binding

When a large number of atoms having only a few weakly bound electrons in the outermost subshell, are brought together, a metallic crystal is formed where the electrons are detached from the atoms and these electrons move freely through the lattice of positive ions. A metal is therefore a regular array of ions embedded in a nearly uniform distribution of free electrons pervading the whole crystal (Fig 16.12). This freedom of movement of electrons in the solid makes them behave in some ways like a gas of electrons. The behaviour of the electron gas is quite different from the ordinary gas of atoms because here the constituents are negatively charged electrons compressed into a much smaller space than the ordinary gas, say in copper there are more than 10^{22} electrons in a cubic centimetre. The electrostatic attraction between the electrons and ions dominates over the mutual repulsion between electrons and between ions, and this is what holds the metal together. The electron gas, therefore, serves the purpose of glue in metallic binding There is no directionality in the binding of the metal. It is for this reason that quite a few metals have close packed structure It is also because of this non-directionality that the metals can be alloyed with one another

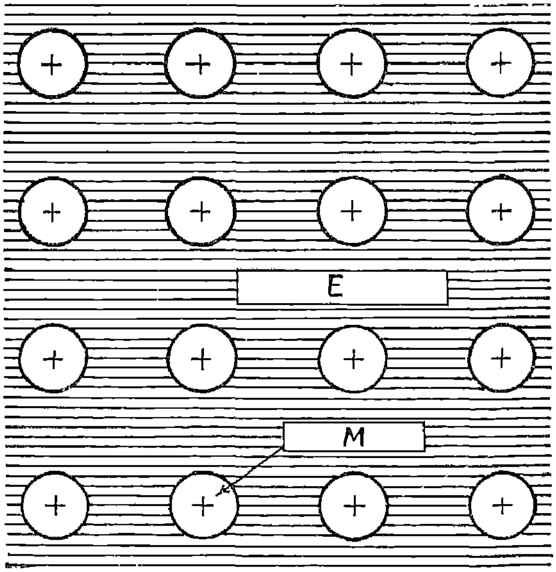

Fig 16 12 A metal can be regarded as an array of ions pervaded by electron gas, E-electron gas, M-metal ion.

The freedom of the movement of electrons through the crystal is responsible for the high electrical conductivity of metals The binding energy of alkali metals, such as sodium, is smaller than in ionic crystals such as NaCl, and such metals are quite deformable.

Van der Waals Binding

Inert gas atoms have completely filled outer shells and these should have virtually no interaction with other identical inert gas atoms It is found that at low enough temperatures (Ne : 2.4K, Ar : 84K, Kr : 117K) inert gases also condense into solids, indicating that there exists a weak attractive interaction between the atoms If there were no such interactions to bind the atoms the system would not condense. In Unit 12, it has been mentioned that inert gas atoms and molecules interact through Van der Waals forces. The bonding in inert gas crystals is due to Van der Waals bonds. In addition to inert gas crystals, the molecular crystals like paraffin, benzene etc., are also held together by the Van der Waals interaction. The Van der Waals bond is considerably weaker than the ionic, covalent and metallic bounds Crystals with Van der Waals binding would therefore have low melting points

From our discussion of the four types of binding in solids we have seen that the interatomic bounds in covalent crystals are directional, and generally the metallic, ionic and Van der Waals bonds are nondirectional. The solids in which the atoms are bonded by nondirectional bonds tend to crystallize in close packed structures. The covalently bonded solids crystallize in structures which conform to the requisite angles between the directed bonds.

16.7 Electrons in Solids

Energy Bands in Solids

In order to understand physical properties of solids we need understand the electronic structure of solids as discussed below.

From our knowledge of the electronic structure of hydrogen atoms, we can say that the energies of the electrons in an atom cannot assume arbitrary values but only some definite prescribed values. Let us illustrate the problem by taking sodium atom which has 11 electrons (Fig 16 13) arranged in the four shells ($1s^2 2s^2 2p^6 3s^1$). The three inner shells (1s, 2s and 2p) are completely filled and there is only one electron in the outermost (valence) shell 3s (which can accomodate 2 electrons when completely filled). The energy levels which are occupied by electrons in the various (1s, 2s, 2p, 3s) shells as well as the next unoccupied shell are shown in Fig 16 14(a). Fig. 16.14(b) illustrates how energy bands arise in solid sodium. As a large number of atoms are brought together, they begin to influence each other Because of the interatomic interaction the electrons in outer shells of the atoms are forced to have energies different from those in atoms. Since a crystal would consist of a large number of the atoms (typically about 10^{23}), we have an enormously large number of levels to deal with. Each atomic level will split into about 10^{23} levels. Because of the such large number of levels in a small energy range (say a few eV at the most) one regards these energy levels to be continuously distributed in energy within a range of energy. We call this an allowed band of energy. These *energy bands* are separated by regions of energy in which no levels exist. This forbidden region of energy is known as band gap or energy gap In Fig 16.14(b) we show how the bands are formed when atoms are imagined to come closer and closer from infinite separation in a crystal of sodium According to the Pauli exclusion principle one level can accomodate only two electrons. If we start filling up nearly continuously distributed energy levels in the bands, we find that the

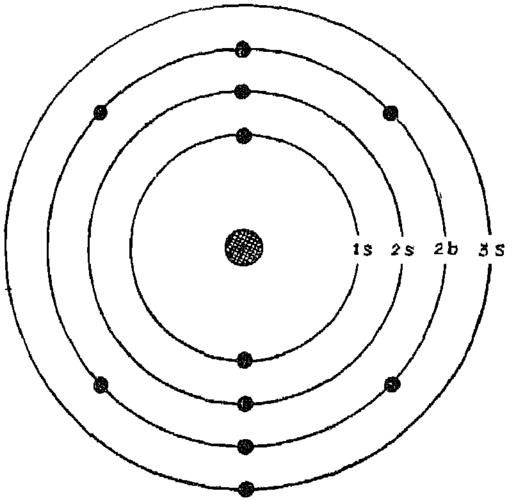

Fig. 16.13. A schematic representation of the electron configuration of the sodium atom.

bond formed out of 3s level will be half filled. This is shown in the Fig 16.14(b) by filling up the occupied portion of the band. The equilibrium separation of atoms in the solid sodium is 3 6 A°. The conventional energy level diagram for solid sodium is obtained from energy levels as they appear at interatomic separation of 3.6 A° in Fig. 16 14(b). Fig 16.15 shows the energy band of the sodium crystal corresponding to the lattice spacing of 3.6 A° (shown by the dashed vertical line on Fig 16 14(b))

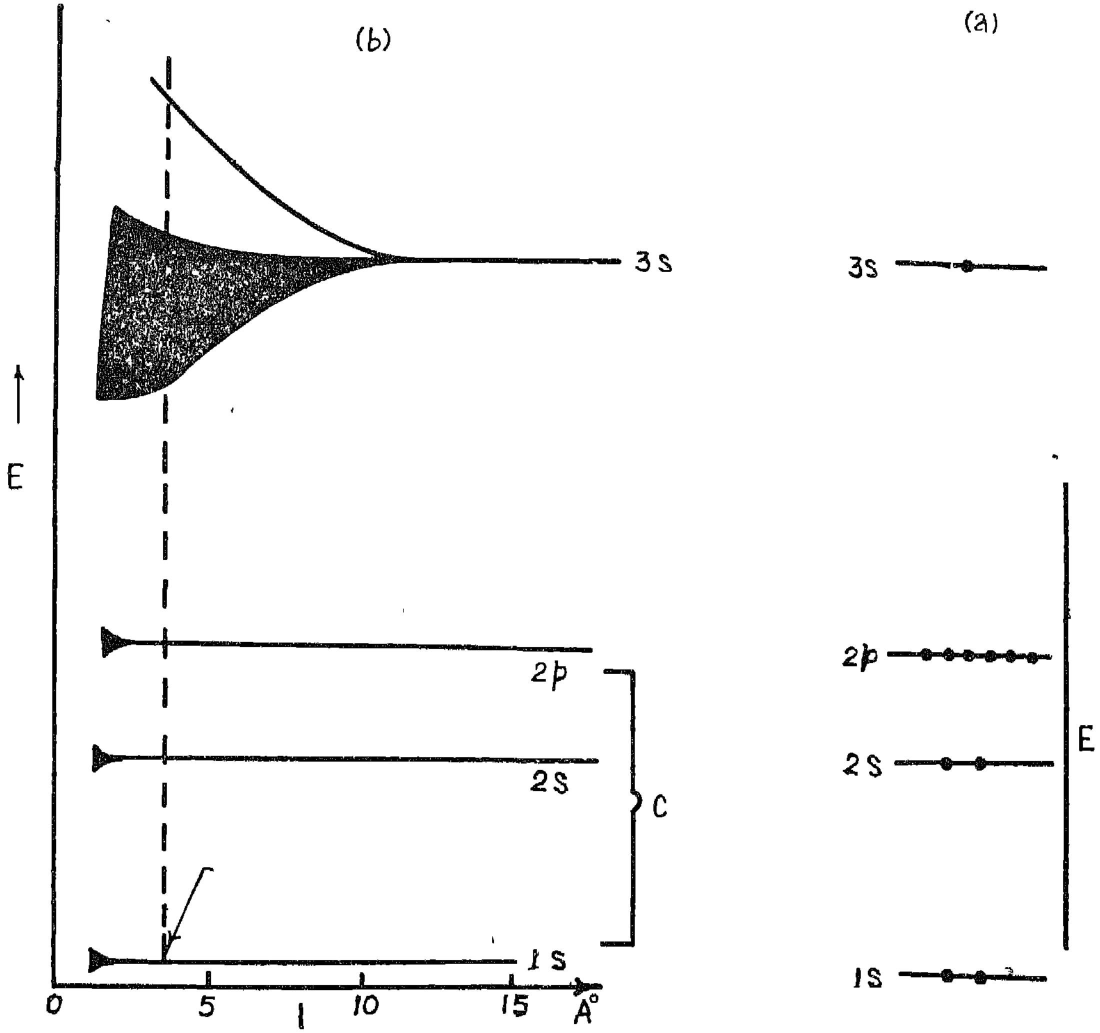

Fig 16 14. (a) Some energy levels of Na atom (b) This schematically shows the formation of energy bands from energy levels of the sodium atoms which are made to approach each other. The dashed line indicates the interatomic seperation in solid sodium E-energy, I-interatomic spacing, C-Ion core states not split in the crystal.

Metals, Insulators and Semiconductors

The band picture for energy states of electrons developed above can be used to classify solids into metals, insulators and semiconductors

Let us consider three possibilities regarding the spacing and the occupancy of the energy bands as shown in Fig 16.16. In case of a metal the topmost band with electrons is not completely filled by them. This partially occupied band is known as the conduction band. In the case of an insulator the uppermost energy band having electrons is completely filled and the next one completely empty but separated from the filled band by a large energy gap. In the case of a semiconductor this gap is comparatively small (say less than about 2eV) silicon with a band gap of about 1eV is a semiconductor where as the diamond with a band gap of 7eV is an insulator. The completely filled band is known as the valence band. The next band which is empty but which has electrons excited from filled states (in valence band or due to impurities) is known as the conduction band.

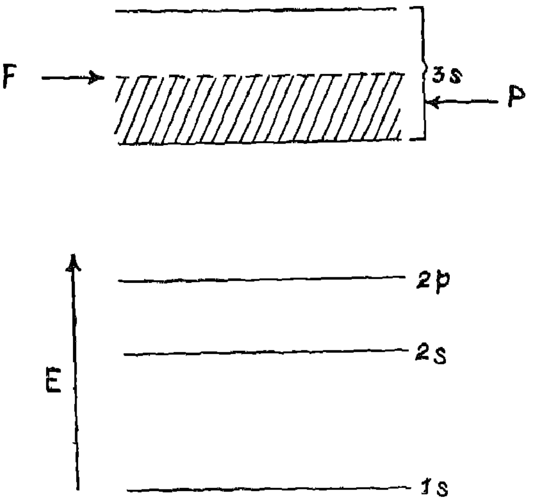

Fig. 16 15 Energy bands for solid sodium Core levels 1s, 2s and 2p are not broadened by interatomic interaction
F = fermi energy, P-partially filled (conduction) band, E-energy.

The electric current flowing through a solid is caused by the momentum gained by electrons due to force acting on them in an applied electric field This process is not possible if the bands in a solid are completely full because the exclusion principle does not permit any redistribution of electrons in a filled band. Electrons, therefore, cannot gain any energy from the field and the current cannot flow. The solid is an insulator In a metal, the conduction band is partially filled. The highest energy level in the conduction band occupied by electrons at absolute zero is called the Fermi level. The energy corresponding to the Fermi level is the *Fermi energy*. In the presence of an applied field the electrons in a metal can be excited to the empty energy levels immediately above the Fermi level and some of the electrons now are accelerated in the direction of field.

At finite temperatures, it is possible for some electrons in the valence band of an insulator with a relatively small gap, to gain enough thermal energy to jump over the energy gap into the conduction band. The probability of the promotion of electrons from valence to conduction band increases exponentially with temperature and decreases also exponentially with the magnitude of the band gap. When a few electrons are excited thermally into the conduction band they leave behind an equal number of empty states near the top of the valence band After promotion of electrons the valence band is not completely full and can contribute to conduction. It is more convenient

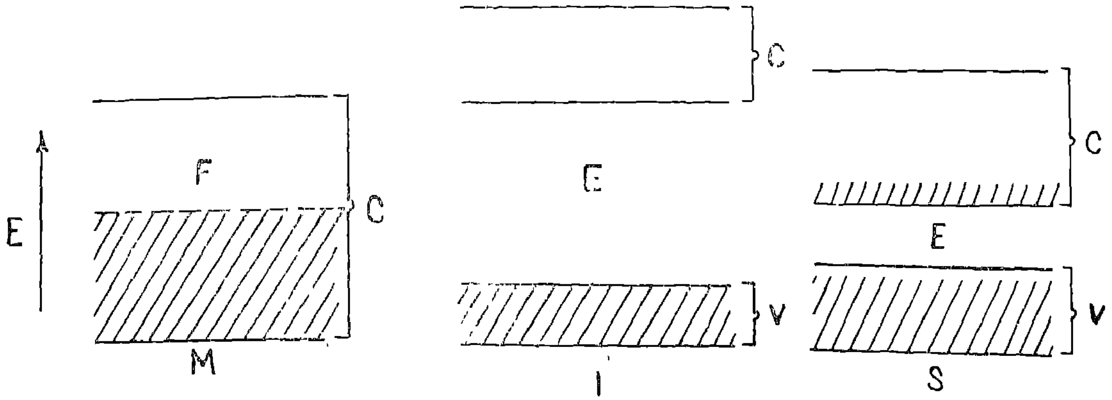

Fig 16.16 Filling of the energy bands in a metal, an insulator and a semiconductor, shaded areas indicate energy regions occupied by electrons E-energy, M-metal, I-insulator, S-semiconductor, F-fermi level, G-energy gap, C-conduction band, V-Valence band

to talk in terms of the few empty states in the valence band rather than the more numerous occupied states These empty states or 'holes' act as if they were positively charged carriers. In semiconductors we shall have electrons in the conduction band and an equal number of holes in the valence band as charge carriers for the current flow.

16.8 Semiconductors

Intrinsic Semiconductors

We have seen above that semiconductors are materials with a small band gap between the valence and the conduction band. The study of semiconductors has led to dramatic developments in electrical devices such as transistors and has brought about a technological revolution in electronics. (See Unit 20).

If we are dealing with a pure semiconductor, its electronic properties are dominated by electrons excited from the valence band (Fig. 16 16). The electrical conductivity in these pure semiconductors is called intrinsic conductivity. In a pure semiconductor at absolute zero the valence band is completely full and there are no electrons at all in the conduction band. At ordinary temperatures appreciable number of electrons would be excited across the energy gap because of the energy of thermal motion of the atoms of the semiconductors. If the energy gap is narrow enough or temperature is high enough, a large number of electrons would be promoted to the conduction band.

We can discuss this intrinsic conductivity in terms of the *bond* model illustrated in Fig. 16.11. Let us take the example of the semiconductor silicon. We had seen in article 16 6, while discussing the covalent bond that the four valence electrons belonging to each silicon atom are used in forming the covalent bonds with its four nearest neighbours. These electrons forming the bonds are not affected when an electric field is applied to the crystal. To say the same thing in terms of the band picture one would say that the electrons forming the bonds completely fill the valence band of the crystal. Substantial amount of energy is needed to release these electrons from their bonding state to a free state in the conduction band and this energy could be imparted thermally due to the heat motion of atoms at any finite temperature. Therefore at any temperature, other than absolute zero, there would always be a finite number of free electrons to carry current and the number of such carriers will rise very rapidly with the increase of the temperature In terms of the language of the band theory this corresponds to thermal excitation of electrons from the valence band to the conduction band A schematic representation of the intrinsic semiconductors is given in Fig 16 17. When an election is broken apart from a bond, a vacancy is created in the bond. This absence of an electron from the bond is called the hole An electron from a neighbouring bond can move into this hole and exchange places with it. The motion of the hole in a particular direction is, therefore, equivalent to the motion of an electron in the opposite direction In this manner in the presence of a field both the electrons and holes can move and give rise to an electric current. The electrons would move towards the positive potential and the holes in the opposite direction.

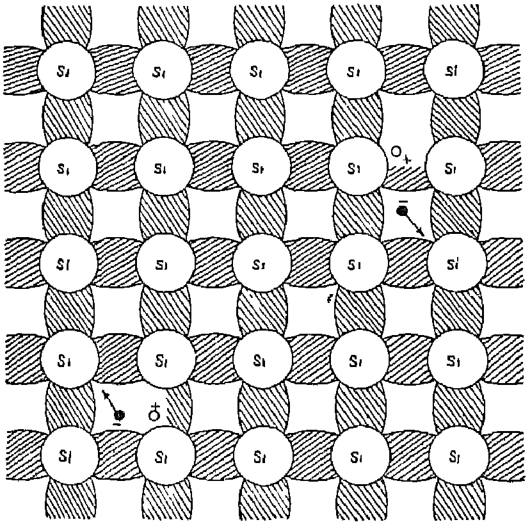

Fig. 16 17 In a pure crystal of the silicon, the electron breaks away due to thermal energy from the covalent bond providing us with an electron free to conduct electricity. This electron leaves behind a vacant state called hole which also move from one bond to another. In the language of the bond theory the electron is promoted thermally from the valence band to conduction band leaving behind a hole.

Extrinsic Semiconductors

Addition of impurity atoms, whose valence is different from four, also introduce electrons and holes into the host crystal and drastically influence the electrical properties of the semiconductor. Atoms of an element belong to group V of the periodic table, say phosphorous or arsenic, can be substitutionally dissolved in Si or Ge Only four of the five electrons of the impurity atoms participate in forming the bonds and the fifth electron from each impurity is almost free to conduct electricity. We say almost free because a very weak attraction of this electron towards the positively charged impurity atoms is always there The coulomb attraction between the positively charged impurity ion and the electron is weakened by the (dielectric constant of the) medium. This would mean that these electrons from impurity atoms would have energies a little less than the energies of the mobile (free) electrons belonging to the conduction band In the band picture the energy states for these electrons would lie a little below the bottom of the conduction band as shown in Fig. 16 18. This electron can be readily freed by energy of heat motion. The impurity atoms which can give up an electron in this manner are called donors. A deliberate addition of

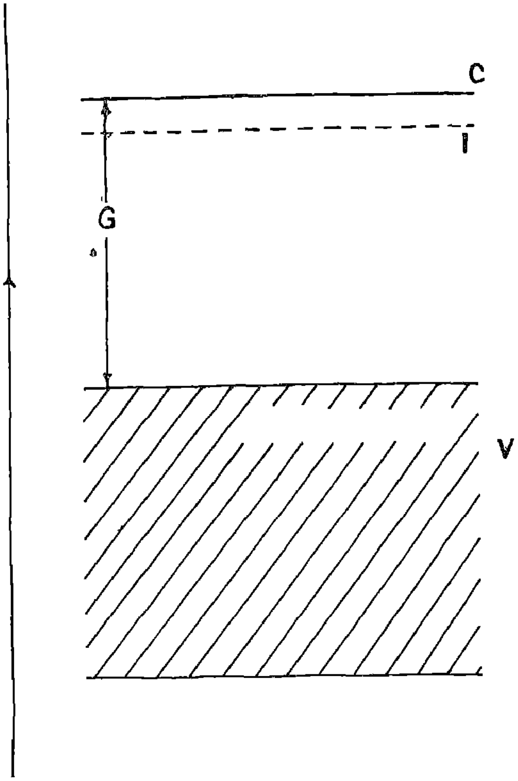

Fig 16 18 Impurity states due to donors in an extrinsic semiconductor.
C-conduction band, I-impurity levels for donors, V-valence band, G-energy gap.

impurities is called doping A semiconductor like Si or Ge doped with pentavalent impurities like phosphorous or arsenic is called an n-type semiconductor because it has an excess of negatively charged electrons n-type silicon is shown in Fig 16 19. The electrical conductivity of the n-type semiconductor will be determined by number of free electrons in the conduction band and the case with which they move in the crystal is called the mobility of electrons.

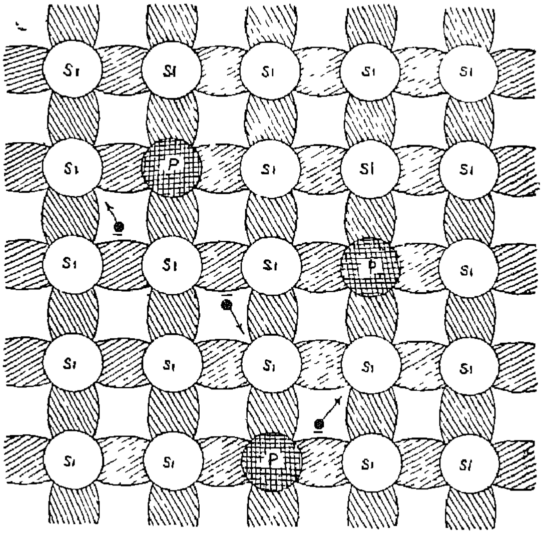

Fig 16 19 Introduction of electron as carrier in a n-type semiconductor obtained by addition of phosphorus (P) to silicon (Si)

If an impurity atom belonging to group III, boron, aluminium, gallium and indium, substitutes an atom of the host semiconductor crystal like Ge or Si, then there would be a vacant place or a hole in the bond because the impurity atom has only three valence electrons. The impurity atoms that give rise to such holes are known as acceptors The hole formed near the impurity may be filled up by an electron from one of the neighbouring bonds and would leave a hole in this bond. This hole may again be filled up by another neighbouring electron. In this manner the hole moves through the crystal. The movement of this deficiency of a negative charge can be regarded as the movement of a positive electric charge. One, therefore, associates a positive charge with the hole. A semiconductor crystal doped with an acceptor is called a p-type semiconductor because it has an excess of positively charged holes. The formatian of holes in p-type silicon doped with boron is

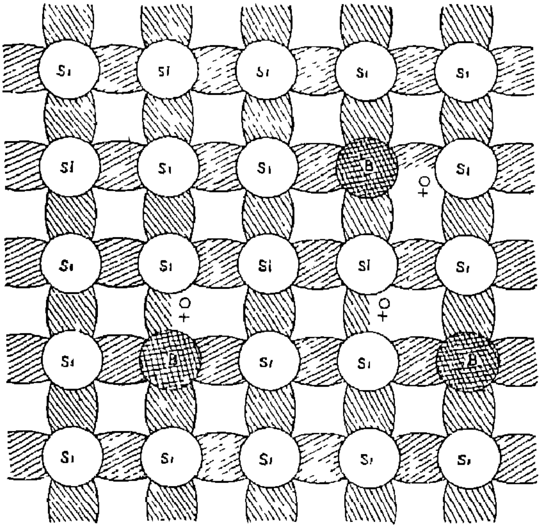

Fig 16.20 A p-type semiconductor can be obtained by doping Silicon(Si) with traces of boron(B). Boron with only three valence electrons to contribute, creates a hole into which a nearby electron can move

shown in Fig. 16.20 Impurity concentrations as small as 100 parts per million lead to an extrinsic behaviour for the semiconductor

The above discussion tells us that the electrical conductivity of both the intrinsic and extrinsic semiconductors would increase with temperature In the case of metals the conductivity decreases with increase of temperature. Irrespective of the manner we introduce electrons or holes in a semiconductor, whether by heating or by adding impurities, these carriers are nearly as mobile as the electrons in a metal But the number of carriers in a semiconductor is much less compared to the electrons in metal, therefore the conductivity of semiconductors is smaller than that of metals.

Exercises

16.1 Calculate the $\frac{c}{a}$ ratio for the hexagonal close packing of spheres with the help of Fig. 16.7.

16.2 Show that the fraction of total volume filled by spheres for hcp structure is 0.74

16.3 Draw the unit cell for the simple cubic (Sc) structure. Find the number of nearest neighbours (6)

16.4 Copper has fcc structure with lattice parameter a=3.6 A°. Estimate the atomic radius of Cu. (1.27 A°)

16.5 Separate out from the following list the materials which form a close packed structure Cu, Zn, Si, Ge, Al, Fe, Ni, Pb, Cd, Ne, Ar.

16.6 How many lattice points are there in the cubic unit cell of the bcc structure? (2)

16.7 What type of bonding do you expect in the following solids :
MgO, NaI, SiC, GiTe, C_6H_6, BN, CuNi.

16.8 Why are metallic solids always opaque, covalent solids sometimes opaque and ionic crystals hardly ever opaque to visible radiation ?

16.9 Why does metallic binding occur mostly with atoms having small number of valence electrons ?

16.10 Of the ionic, covalent, metallic and Van der Waals solids which class is most likely to produce an insulator, a conductur, a semiconductor ?

16.11 What elements other than As and Sb can be used as an impurity with germanium to form an n-type semiconductor ? What elements other than gallium and indium can be used to form a p-type semiconductor ?

16.12 Indicate the nature of binding in the solids with the following description :

(a) Reflects light in the visible, electrical resistivity increases with temperature, melting point 1000°C.

(b) Reflects light in the visible, electrical resistivity decreases with temperature and melting point above 1000°C.

(c) Transmits light in the visible, conducts electricity only at high temperatures ; and

(d) Transmits light in the visible but very low melting point.

16.13 One alloys copper with zinc, will the band width of the conduction band increase or decrease ?

16.14 What factors determine the electrical conductivity of a metal ?

UNIT 17

Properties of Materials

17.1 Elastic Properties of Solids

A lot of our technological development has been possible through our understanding of the behaviour of materials. Our search for materials possessing desired mechanical properties has met with partial success Methods have been developed by which we can improve the mechanical properties like hardness, strength, brittleness, etc., of alloys ceramics and polymers; but we still have not reached the stage where we can make materials having a desired value for a mechanical property.

The study of mechanical properties of a solid in general deals with the deformations due to forces applied on the materials. A solid tries to maintain its shape under the influence of small external forces and recovers its original shape and size when the forces are removed. This property of the solid is known as elasticity.

Young's Modulus

If we stretch a material with certain force in a particular direction, then the material gets elongated along the direction of the force as shown in Fig. 17.1.

As a result of external forces there would arise internal forces within the material and stress T refers to these internal forces. The *Stress* T is defined as force per unit cross-sectional area The *strain*, e, is a measure of the deformation of the material.

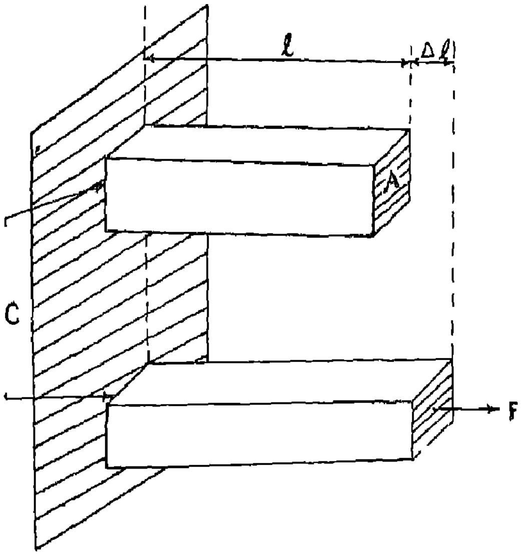

Fig. 17.1 Elongation of a bar under a tensile stress. C—clamped end

When a body is strecthed the strain is defined as the elongation per unit length. If a force F acts on a bar of length l and cross sectional area A, and produces an elongation Δl, then

$$\text{Stress} = T = \frac{F}{A}$$

$$\text{and strain} = e = \frac{\Delta l}{l}$$

Results of experiments show that the relationship between the stress and strain depends on the material The stress-strain relationship for a typical material like copper is schematically shown in the Fig. 17.2. In the limit of very small elongations the stress is proportional to strain.

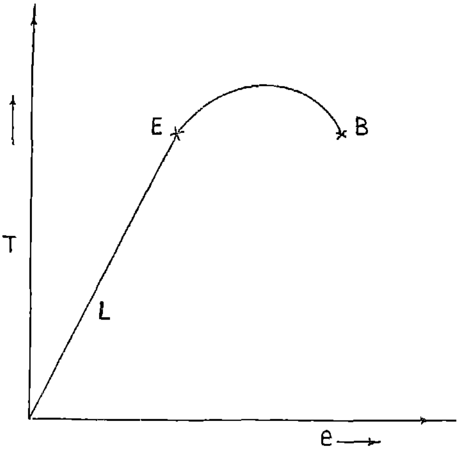

Fig. 17.2 Stress [T] and strain [e] graphs for a material like copper which undergoes deformation before breaking L—linear region, E—elastic limit, P—point at which the material breaks

$$T = Ye \qquad \ldots\ldots(17.1)$$

The constant of proportionality Y is known as *Young's modulus* Y depends only on the nature of the material. The linear relationship 17.1 between stress and strain is known *as Hooke's law*. A low value of the Young's modulus means a material which stretches a lot for small stresses

EXAMPLE 17.1

How much will a 3.0 m long copper wire elongate if a weight of 10 kg is suspended from one end and the other end is fixed. The diameter of the wire is 0 4 mm.

Solution

We shall take Y for copper to be 1×10^{11} N/m^2 and $g = 9.8$ m/sec^2.

$$\therefore \text{ stress} = \frac{mg}{\pi r^2} = \frac{10 \times 9.8\text{ N}}{\pi(2\times10^{-4})^2\text{m}^2} = 7.8 \times 10^8 \text{ N/m}^2$$

$$\text{strain} = \frac{\Delta l}{l} = \frac{\Delta l}{3}$$

$$\text{Hence} \quad \frac{\Delta l}{3} \times (1\times10^{11}) = 7.8\times10^8$$

$$\text{or } \Delta l = 2.34 \text{ cm.}$$

Non-elastic Behaviour

If a specimen returns to its original shape and size on removal of the stress, the deformation is said to be within the elastic limit. Beyond the elastic limit, the applied forces produce plastic deformation and the body will not recover its orginal shape even if the stress is withdrawn i e. the material undergoes a permanent deformation. If we increase the strain beyond the elastic limit the material will break at some value of strain. A material which breaks in this way is called ductile. A ductile material is therefore permanently distorted before breaking. Most of the metals are ductile.

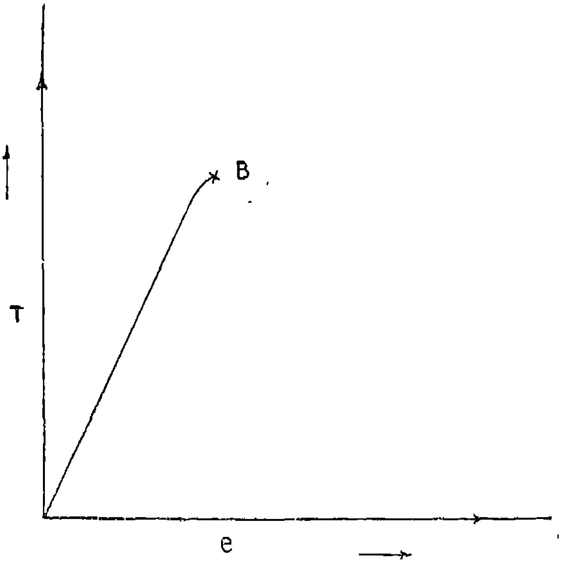

Fig. 17.3 Stress [T]—Strain [e] curve for a brittle material like cast iron. B—breaking point,

Some metals like cast iron or materials like glasses and ceramics break quite suddenly as soon as the stress-strain curve starts deviating from a straight line (Fig 17.3). Such materials are called *brittle.* Brittle materials are difficult to form into different shapes where as ductile materials can be shaped.

In the above we have only considered the tensile (or compressive) force which stretches (or compresses) a solid. We can also deform a solid by hydrostatic stresses that is by pressures or tensions acting uniformly over the whole surface of the body or else by twisting forces which we will discuss below.

Bulk Modulus

If a body is subjected to uniform pressure acting normal to its entire surface it is said to be subjected to 'hydrostatic' pressure' (Fig. 17.4). A body submerged in liquid will feel such a pressure. Even though such pressures may be exerted by any other method, the pressure is called hydrostatic. As shown in the Figure if the pressure (which is the force per unit area), P, on the body is increased its shape remains unchanged, but the volume changes. The strain is defined as the ratio of the change in volume $\triangle V$ to the original volume. The stress within the body is equal to the change in pressure $\triangle P$,

$$\therefore \text{Stress} = K \times \text{strain}$$

where the constant of proportionality K is known as the bulk modulus of elasticity. Therefore

$$K = -\frac{\triangle P}{\triangle V/V} = -V\frac{\triangle P}{\triangle V} \quad \dots 17.2$$

The minus sign is included because an increase in the hydrostatic pressure leads to decrease in the volume, that is if $\triangle P$ is positive, $\triangle V$ is negative

The reciprocal of the bulk modulus is known as the compressibility.

Shear Modulus

If a pair of forces G act in opposite directions upon the upper and lower faces of a rectangular block, these forces shear it into a parallelopiped, turning the vertical and faces through small angle (Fig. 17.5). The shear stress is defined as the tangential force on a face divided by the area A of the face. The shear strain is the angle of twist. μ known as the shear modulus or coefficient of rigidity relates the stress and strain.

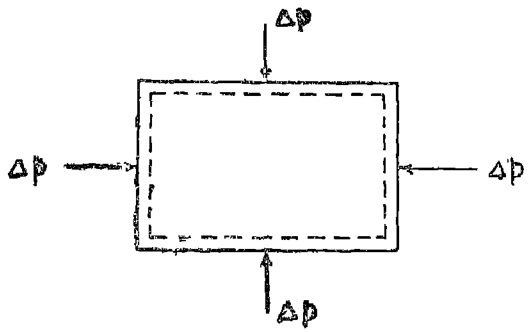

Fig. 17.4. Hydrostatic pressures Δ p causes a decrease in the volume.

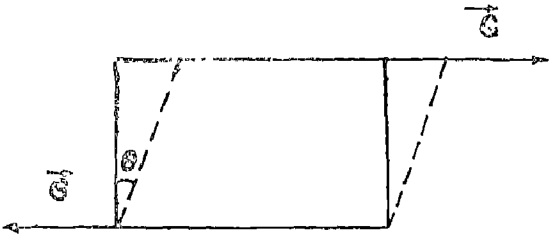

Fig. 17.5 The shear stress produces a change in shape without changing the volume.

$$\text{stress} = \mu \times \text{strain},$$

$$\text{so that } \mu = \frac{G/A}{Q} \quad \dots 17.3$$

17.2 Elastic Properties and Interatomic Forces

We had seen earlier that the atoms in a solid, interact with each other through interatomic forces. If we could plot the potential energy of interatomic interactions against interatomic sepeiation we would find that in a crystal the atoms are located at positions of minimum potential energy. The net force (which is the derivative of the potential) on an atom is zero at these potential minima A tensile stress will increase the separation of atoms along the direction of the stress and the atoms are displaced from the potential minima to positions of higher energy due to the applied force. If the force is released the atoms will move back to their equilibrium positions. The displacement suffered by individual atoms for a particular tensile force will depend on the strength of the interatomic forces Stronger the interatomic forces, smaller will be the displacements (strains) and higher would be the value of the young's modulus.

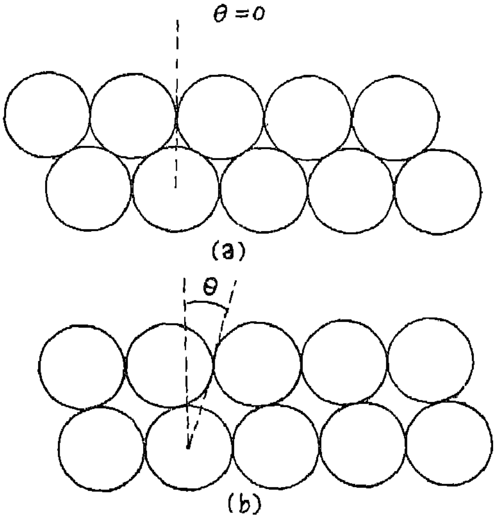

Fig. 17.6 (a) Two adjacent planes of atoms are shown in a stress free crystal.
(b) A shear strain makes these planes slide past each other.

In the case of shear stresses the distance between atoms remains the same, but neighbouring planes of atoms in a solid slide past each other. Figure 17 6 (a) show the initial location of atoms in two neighbouring planes, the location of atoms after the shear deformation is shown in Fig. 17.6 (b). This corresponds to the shear strain shown in the figure.

17.3 Thermal Properties of Solids

In our daily life we need materials which could store heat, materials which expand very little when heated, materials which are bad (or good) conductors of heat. We, therefore, need to understand thermal properties of materials. We discuss here qualitatively the heat capacity, the thermal expansion and the thermal conductivity of solids.

Heat Energy

Heat energy is identified with the motion of atoms in a solid, while describing the structure of solids we assumed that the atoms sit at the lattice sites. But an atom really does not remain fixed at a lattice site, but oscillates about it. A configuration of atoms at an instant is shown in Fig. 17.7. The thermal energy is stored in the vibratory motions of these atoms. Each atom experiences a potential energy of the form shown in Fig. 17.8. In the absence of any kinetic energy, the atoms will sit at their potential minima. This would be the situation at O°K. At a higher temperature the atoms would execute oscillatory motion. In the neighbourhood of the minimum the poten-

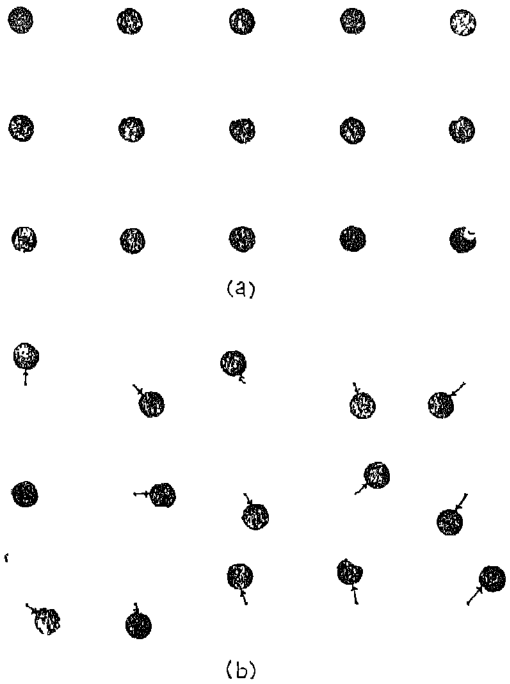

Fig. 17.7 (a) Idealised picture of atoms sitting at lattice sites ; (b) Instantaneous configuration of vibrating atoms in a solid.

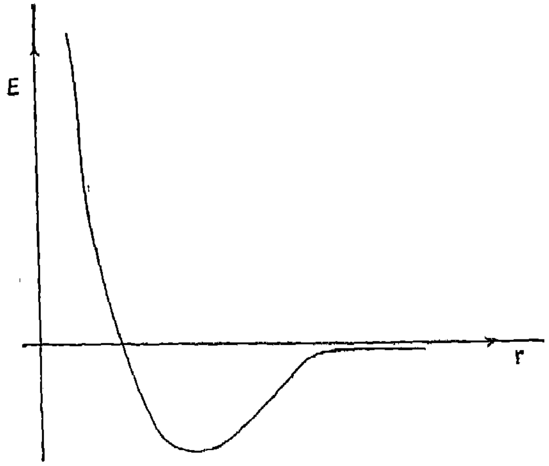

Fig. 17.8 Schematic form of interaction between atoms in a solid ; r—separation between atoms, E—potential energy.

tial function would be nearly parabolic. A parabolic potential energy means that the motion of the atoms would be simple harmonic. But an atom does not oscillate independently of other atoms A displacement of an atom in a particular direction would influence the position of its neighbouring atoms and thus the motions of atoms are correlated. These facts, therefore, suggest a mechanical model for the vibration of atoms The thermal energy of the solid is the energy of the vibrational motion of these atoms. In this model the solid may be regarded as a system of mass points linked by springs (Fig 17 9). The motion of such a system would be quite complicated This energy of the vibratory motion will be greater if the mass points have larger displacements from their equilibrium positions. Under ordinary conditions the amplitude of these atomic oscillations are of the order of 10^{-9} cm and the frequencies of the oscillations are about 10^{13} HZ.

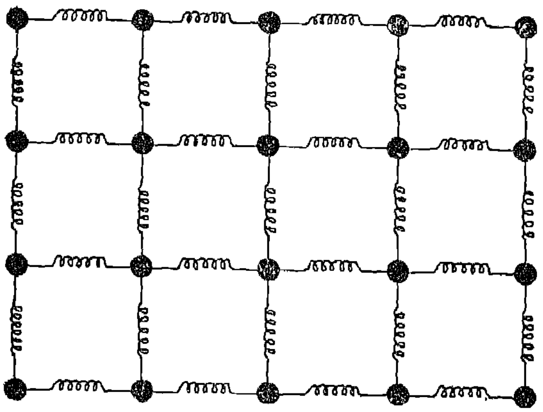

Fig. 17.9 A mechanical model for a vibrating crystal lattice

Specific Heat

We have seen above that when heat energy is added to a solid, it is used up in exciting the vibratory motion of the atoms and the temperature of the solid rises. If the quantity of heat energy $\triangle Q$ supplied to a standard mass of a substance under certain conditions leads to an increase in temperature ΔT, then we define $\Delta Q/\Delta T$ as the specific heat of that mass. If one chooses the standard mass to be a mole (also called gram molecule) the specific heat is called the molar specific heat The units of the molar specific heat are J mol^{-1} deg^{-1}. Experimental measurements show that near room temprature the value of specific heat of nearly all solids is close to 3R or 25 J mol^{-1} deg^{-1} where R is the gas constant. This is known as the Dulong and Petit rule. At lower temperatures the specific heat decreases rapidly as shown in Fig. 17.10.

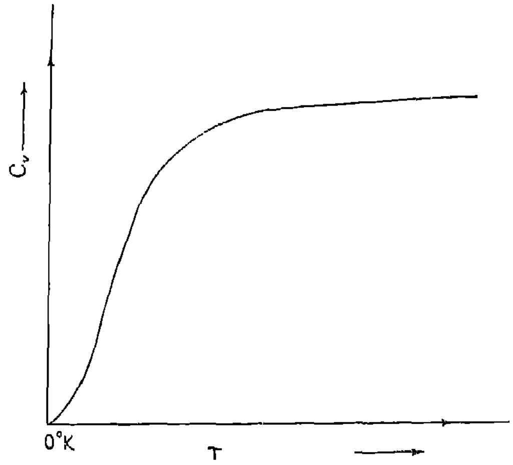

Fig. 17.10 The specific heat of a typical solid at constant volume C_V as a function of temperature, T

It is not possible to explain the drop in the value of the specific heat with the lowering of temperature using a classical theory. The explanation by Einstein of the variation of the specific heat with the temperature as well as with the nature of the solid was one of the earliest successes of the quantum theory. We shall not discuss the details of it here

17.4 Thermal Expansion

Nearly all solids expand on heating. The change in length (or width or thickness) of a solid is called a *linear expansion* If the initial length is l and a change ΔT in temperature leads to a change Δl in the length, then we find from experience that Δl is proportional to ΔT as well as l. Hence we are able to write

$$\Delta l = \alpha l \Delta T$$

$$\text{so/that} \quad \alpha = \frac{\Delta l}{l \Delta T} \qquad \text{... (17.4)}$$

where α is known as the coefficient of linear expansion.

Because the length, width and thickness of a solid increase with temperature, its volume will also increase. The change in volume ΔV of a solid on increasing its temperature by ΔT is proportional to the initial volume V and T Therefore we can write

$$\Delta V = \beta V \Delta T$$

$$\text{and} \quad \beta = \Delta V / V \Delta T \qquad \text{... (17.5)}$$

where β represents the volume coefficient of expansion. We assume the solid to be isotropic for which the coefficient of linear expansion is the same along any direction. For example, if we take a rectangular block of the solid as shown in Fig. 17.11, the coefficient of linear expansion α will be the

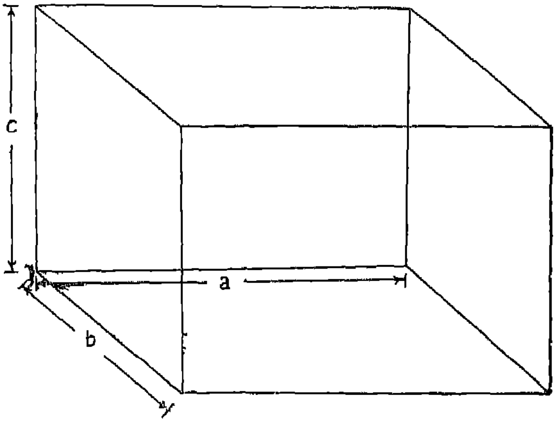

Fig. 17.11 A rectangular block of a material.

same for the length, the width and the thickness or for that matter, for any other length in this block. One can show that for an isotropic solid the volume coefficient is three times the linear coefficient of expansion. In order to show this, consider the rectangular block of Fig. 17.11 whose length, width and thickness are given by a, b and c respectively. The volume is therefore

$$V = abc$$

If we increase the temperature of the block by 1K, the changed length, width and thickness would become $a(1+\alpha)$, $b(1+\alpha)$ and $c(1+\alpha)$. Therefore the volume is

$$V + \Delta V = a(1+\alpha)\ b(1+\alpha)\ c\ (1+\alpha)$$
$$abc\,(1+3\alpha+3\alpha^2 + \alpha^3)$$

Because the coefficient of thermal expansion α is always a small quantity, we can neglect $3\alpha^2 + \alpha^3$ and write

$$V + \Delta V = abc + abc\,3\alpha$$

Because $V = abc$, we can write for change in volume for one degree rise in temperature to be

$$\Delta V = 3\alpha V$$

This by definition is equal to β V, therefore

$$\beta = 3\alpha \qquad \dots (17.6)$$

In terms of our atomic picture, thermal expansion is due to an increase in the average distance between atoms as the temperature increases. As we increase the temperature the heat energy available for atomic oscillations increases. This would mean that with the rise in temperature the amplitude of the ocillation of atoms would increase. If the} interatomic potential energy is exactly symmetrical with respect to the point where the potential energy is minimum, we see from Fig. 17.12 that the average interatomic distance between atoms does not change when the amplitude of atomic vibrations increases with the rise of temperature. The atom vibrates about the same average interatomic sepration with a

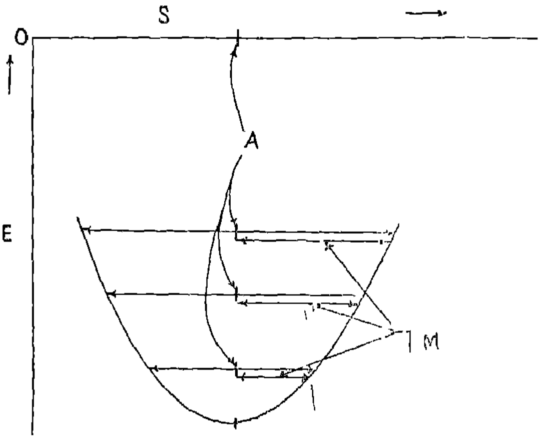

Fig. 17.12 Vibrational motion of atoms in a solid for a symmetrical interatomic potential energy (E). S—interatomic separation ; A—average inter-atomic sepation ; M—amplitude of vibrational motion.

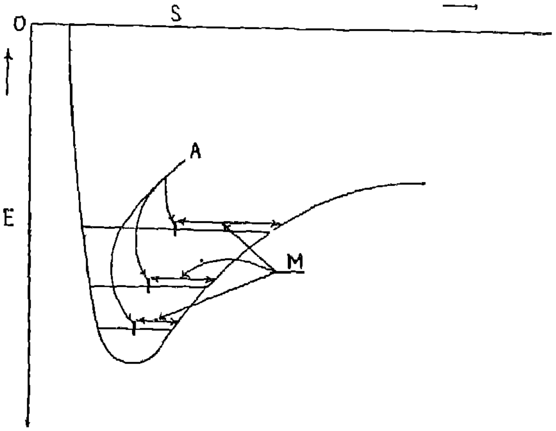

Fig. 17.13 The interatomic potential (E) for a real crystal is asymmetric. S—interatomic separation , A—average interatomic separation ; M—amplitude of vibrational motion.

larger amplitude at a higher temperature. We see from Fig. 17.12 that thermal expansion would not exist if we use a symmetric potential

The actual interatomic potential is somewhat asymmetric (not symmetrical). We have shown such an asymmetric potential in Fig 17 13. We see in the figure that as the amplitude of oscillation increases with temperature, the average interatomic separation also increases. We can also see that the value of the thermal expansion coefficient will depend on the details of the shape of the potential energy function In general the more asymmetric the potential, the larger would be the value of the thermal expansion coefficient.

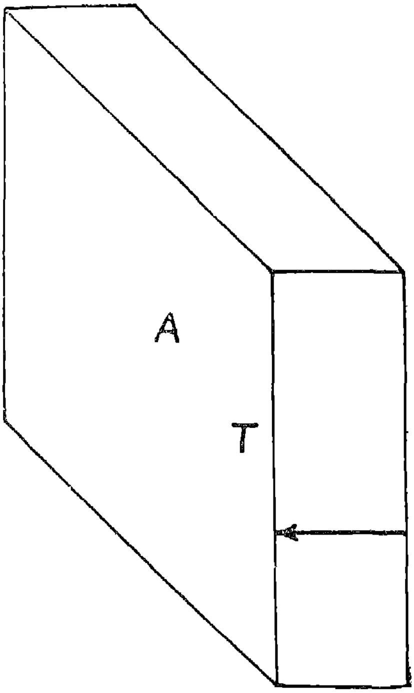

Fig. 17.14 Heat flow through a slab.

17.5 Thermal conductivity

If we take a piece of a solid and heat one end of it, then heat energy will flow from the hot end to the cold end. It is because of this reason that when we place one end of a metal rod over a flame, we feel the other end also heated up The phenomenon of transfer of heat energy from one part of a solid to another part at a lower temperature is known as the conduction of heat

Consider a slab of material of thickness Δx and cross sectional area A (Fig 17 14). The two faces of the slab are at temperatures T and T + ΔT An amount of heat ΔQ will flow from the hotter face of the slab to the colder one in time Δt. Experimentally one finds that the rate of flow of heat through the slab, $\Delta Q/\Delta t$ is directly proportional to the temperature difference ΔT, and to area A, but is inversely proportional to the thickness Δx. Therefore

$$\frac{\Delta Q}{\Delta t} \propto A \frac{\Delta T}{\Delta x}$$

The constant of proportionality is represented by K called the thermal conductivity of the material. Thus the heat conduction equation becomes

$$\frac{\Delta Q}{\Delta t} = KA \frac{\Delta T}{\Delta x}$$

A material for which thermal conductivity K is large, is a good conductor like copper and aluminium For thermal insulators like glass, wood and ice K has a low value. Usually metals are good conductors of heat and insulators are poor conductors

On the atomic scale a picture of thermel conductivity of an insulator can be presented as follows When a part of a solid is heated, the atoms there start oscillating violently about their equilibrium position, As mentioned, all atoms are bound to each other through interatomic interactions, therefore oscillation of an atom sets the neighbouring atoms inio oscillation. The atoms at the hotter end will be vibrating with larger amplitudes than atoms at the colder end. An atom vibrating with larger amplitude will try to increase the amplitude of vibration of its neighbour which initially was vibrating with a little smaller amplitude and thus gives energy to it This neighbour in turn give energy in the same way to its neighbour and so on. Thus the atoms at the colder end will receive energy from atoms at the hotter end of the material.

In the case of metals heat energy can also be transported by electrons in addition to the mechanism discussed above for insulators Since the electrons are quite mobile they are much more effective in transporting energy from hotter to colder parts of a solid compared to the mechanism given above. The thermal conductivity of metals is therefore, essentially determined by electrons.

In a metal the electrical conduction is due to the mobile electrons. Since the thermal conductivity of metals is also due to conduction electrons coefficient of the thermal conductivity of a metal is related to its electrical conductivity In an insulator (which we shall also call dielectric) there are no conduction (or free) electrons and we may tend to think that the electric field will have no effect on the insulator. This is not the case. When a dielectric is placed in an electric field, if the molecules constituting the dielectric have a permanent dipole moment, then the dipole moment of the molecules will tend to align themselves along the direction of the field as shown in Fig. 17.15. Thermal motion of molecules does not permit a complete alignment. In the absence of the field these dipoles were randomly oriented. Even if the constituent molecules do not possess a permanent dipole moment, the applied field will induce dipole moments in them. The induced moments would already be lined up along the direction of the field, because the induced dipole moment is along the direction of the field. Therefore, an electric field will polarize a dielectric.

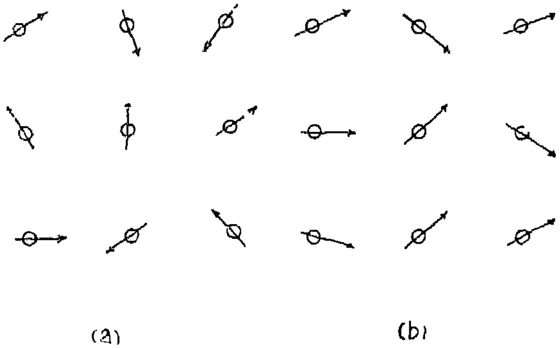

Fig. 17.15 Effect of an electric field on a dielectric whose constituent molecules possess permanent dipole moment.

If we place a slab of a dielectric between the plates of a parallel plate capacitor, the electric field will polarize the dielectric i.e. the negative charge of the molecule will be shifted slightly towards the positively charged plate and the positive charge towards the negatively charged plate (Fig 17 16 a). These displacements are very small and are of the order of molcular diameters (10^{-10} m. In the bulk of the slab there will be no net charge but at the top surface in Fig. 17 16. (b) there would be an excess of negative charge and at the bottom surface there would be an excess positive charge due to polarization of the material. The charges in the capacitor plates set up an electric field E_0, the induced surface charges on the dielectric slab will set up a field—E' in a direction opposite to that of E_0 Therefore the resultant field, is E_0-E'. Thus when a dielectric is placed in an electric field, the surface charges are induced in the dielectric, which tend to reduce the original field This is shown in Fig. 17.16 (c). The ratio $\frac{E_0}{E_0-E'}$ is known as the dielectric constant of the dielectric. If the dielectric consists of polar molecules, their orientation

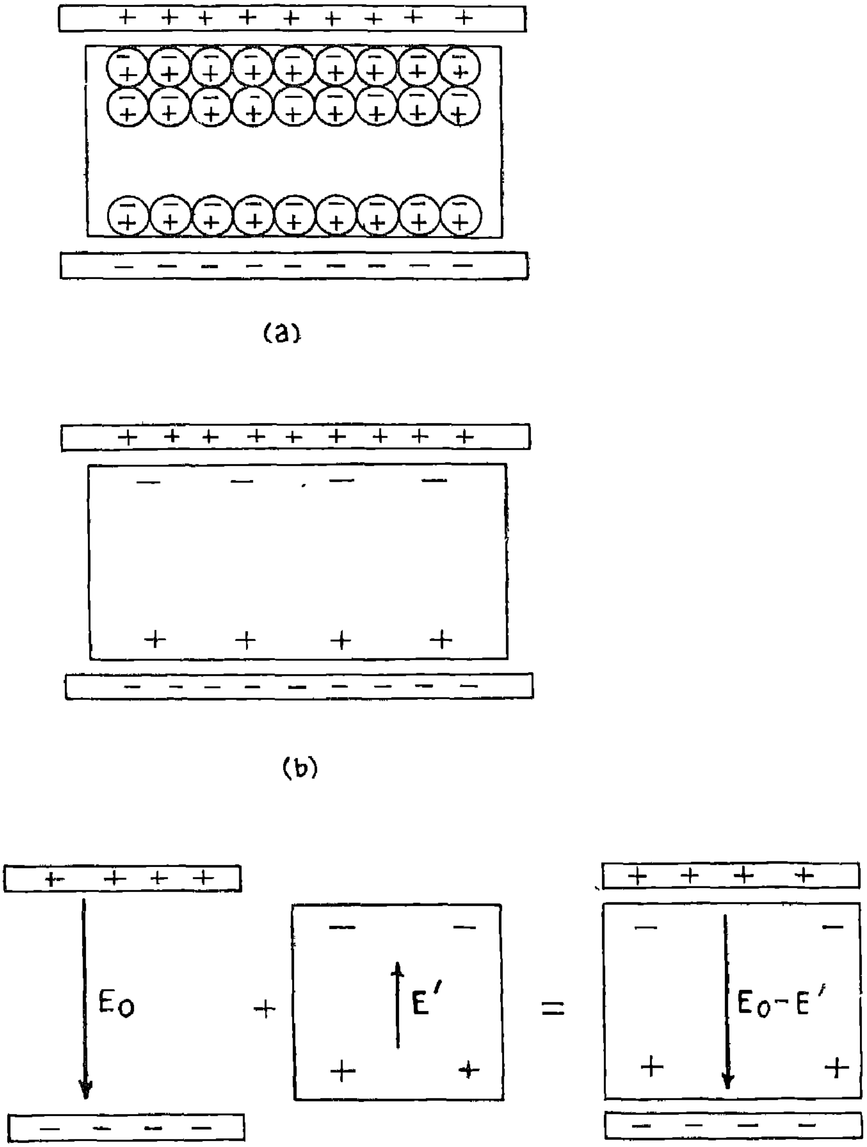

Fig. 17.16 Capacitor with a slab of dielectric placed between the plates.

in the applied field would give a larger valve of E′, therefore a smaller E_0-E′. The dielectric constant of such a dielectric would be large.

17.7 Magnetic Properties of Materials

From the point of view of the magnetic behaviour we can classify the materials into three categories : 1 diamagnetic, 2. paramagnetic, and 3. ferromagnetic. In all materials except the ferromagnetic materials the magnetic effects are quite small. Ferromagnetism is shown by materials like iron, nickel, cobalt. Any magnetic field would be strongly influenced by the presence of ferromagnetic materials. A piece of diamagnetic material (e g. bismuth) when placed in a non-uniform field will experience a force directed from stronger to the weaker part of the field. A piece of bismuth brought near the pole of a strong magnet will be repelled. This will be a weak effect. A paramagnetic matarial, salts of transition metals or rare earths like chrome alumn placed in a non-uniform field will experience a force directed from weaker to the stronger part of the field. Therefore, a piece of aluminium placed near a strong magnet will feel a weak attraction. A ferromagnet will experience a relatively larger force directed from weaker to the stronger part of the field. The effect for a ferromagnet is so large compared to a paramagnet or a diamagnet that paramagnets and diamagnets are popularly called 'nonmagnets'.

The magnetic properties of matter are traced to motion of charges. The motion of an electron in an atom or a molecule would lead to magnetic effects. The spin of the electron is also a source of magnetism of the particle. It is not possible to present an accurate picture of magnetism without using quantum mechanics, because spin of an electron as well as its motion in a solid can be described accurately only by using quantum mechanics. We shall, therefore present models which will give qualitative picture of diamagnetism, paramagnetism and ferromagnetism.

Diamagnetism

We had seen earlier that in an atom the electrons move in circular orbits around the nucleus. The rotating electron will behave like a current loop This current loop is equivalent to a magnetic moment equal to the product of the current and the area of loop (here the electron orbit). The direction of the moment is normal to the plane of the orbit (Fig. 17.17 a). In a material there will be a very large number of such electrons and the orbits of the electrons will be oriented randomly. The magnetic moments due to orbital motion of electrons cancel each other and there is no net magnetism, Fig. (a) and (b). If a magnetic field $\vec{B}$ is impressed from outside, the electron

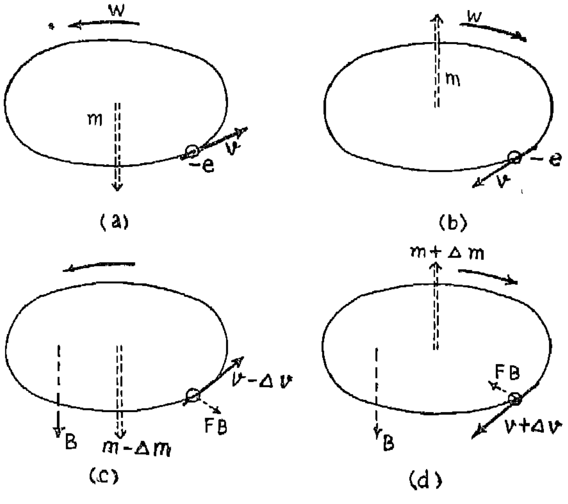

Fig. 17.17 An electron orbiting in an atom produces a moment.

experiences a radial force $\vec{F}_B = -e(\vec{v}\times\vec{B})$ which is at right angles to the direction of motion. Since the radius of the orbit remains constant, the angular velocity of the electron will decrease or increase depending on whether F_B acts radially outwards, Fig (c) or inwards, Fig. (d) Thus the effect of applying the magnetic field is to decrease or increase the angular velocity of the orbiting electron. A reduced angular velocity in (c) would mean, a reduced current and therefore a reduced magnetic moment. Similarly the magnetic moment of the electron orbiting an opposite direction in (d) would increase the magnetic moment. The two electrons orbiting in the manner shown in Fig. (a) and (b) pair up to give a net zero moment in the absence of the magnetic field. In the presence of the field such a cancellation is not possible because the net effect of the two orbits (c) and (d) in Fig. 17.17 is to give a moment in a direction opposite to that of the impressed field. The moments due to all the electrons in a material in the presence of an external magnetic field will add up to a finite magnetization of the material in a direction opposite to that of the field. This accounts for diamagnetism of the material This also shows that diamagnetism should be present in all materials because the orbital motion of electrons in atoms is a universal phenomena.

Paramagnetism

From the above discussion you might expect all materials to be diamagnetic. But experiments show that this is not the case. Many materials are paramagnetic because in them paramagnetism completely dominates the diamagnetism. The paramagnetism is due to the fact that in many solids, the ions or atoms in it have permanent magnetic moment. These moments come from the orbital motion of the electrons and from the magnetic moment intrinsic to each electron arising from its spin. We find that in atoms whose inner electron shells are not completely filled there is a net moment in them. This is the case for atom of transition elements like iron, nickel, cobalt, palladium and for rare earth atoms like gadolinium, etc.

If we place a magnetic moment $\vec{m}$ in a magnetic field $\vec{B}$, the field will try to align the moment along the direction of the field, because a moment pointing along the direction of $\vec{B}$ will have the lowest energy. Let us consider a solid containing N atoms each of which has a magnetic moment $\vec{m}$. The interactions between these moments is very weak and we can regard these moments to be independent of each other. In the absence of a magnetic field these atomic moment would be completely randomly oriented because of the thermal agitation of the atoms at any finite temperature. When we apply an external magnetic field, it will try to line up the magnetic moments along the direction of the magnetic field. This process of alignment along the direction of the magnetic field is hampered by the disorienting influence of the thermal motion of the atoms. As we increase the field, more and more moments will orient themselves along the field, but if we increase the temperature this orientation will be opposed. At high enough field at a temperature T all moments would line up along the direction of the field. This will give us the maximum possible magnetic moment $N\vec{m}$. At low

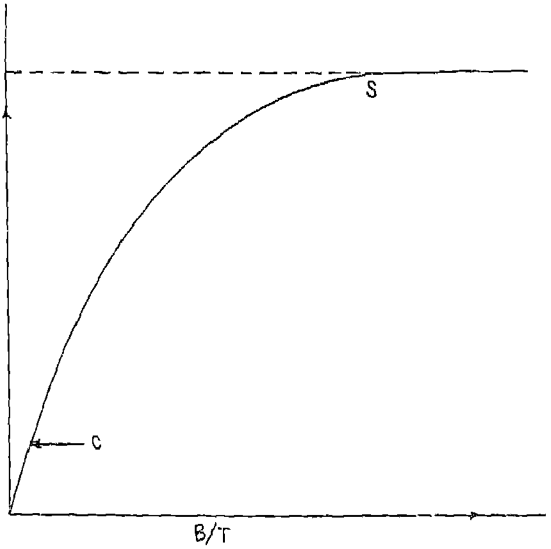

Fig. 17.18 Total magnetic moment (M) of a paramagenetic solid as a function of $\frac{B}{T}$; S—Saturation region, C—Curie's law region

fields the total magnetic moment would be directly proportional to the magnetic field $\vec{B}$ and inversely proportional to temperature T. As one increases the field or lowers the temperature the total moment would reach the saturation value $N\vec{m}$. Fig. 17 18 shows how total magnetic moment varies as a function of $\frac{B}{T}$. At low $\frac{B}{T}$ when we are far from the saturation region we can see that the magnetisation M (which is defined as the total magnetic moment per unit volume) would be given by

$$M = C\frac{B}{T}$$

Where C is a constant. This relation is known as Curie's law. This was discovered by Pierre Curie experimentally.

We thus see that for a solid to be paramagnetic it is necessary that some of its constituent atoms or ions should possess a magnetic moment M If they do not possess a moment then there will be no paramagnetic effect and the material would be diamagnetic. As has been said before the presence or absence of an atomic moment is decided by the electronic structure of the atom or the ion.

Ferromagnetic Materials

The magnetisation observed in ferromagnetic materials is much larger than in paramagnetic substances. The elements iron, cobalt, nickel, gadolinium and dysprosium and a number of alloys are ferromagnetic. In a ferromagnetic material the magnetic moments of the individual atoms are coupled to neighbouring moments very strongly through an interaction which has come to be known as exchange interaction. The exchange interaction in ferromagnetic materials align the moments all in the same direction as sketched in Fig 17 19.

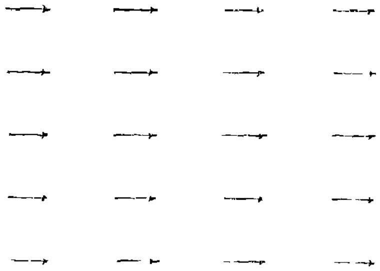

Fig. 17.19 In a ferromagnetic material the exchange interaction aligns all the moments along the same direction.

In a ferromagnet if all the moments are lined up parallel to each other why is it that every piece of iron or nickel is not a magnet? The reason is that an ordinary piece of unmagnetized ferromagnet consists of numerous domains in each of which the moments are pointing along the same direction But in neighbouring domains the moments of a domain are pointing in different directions A piece of a ferromagnet consists of a very large number of domains all pointing in different directions and no direction is preferred over others This means that the total moments due to domains will cancel each other and the piece of the ferromagnet will appear unmagnetized. Fig 17.20 shows the orientations of domains in a ferromagnetic material. In the presence of a magnetic field the domains align themselves along the direction of the field to give a strong magnetization.

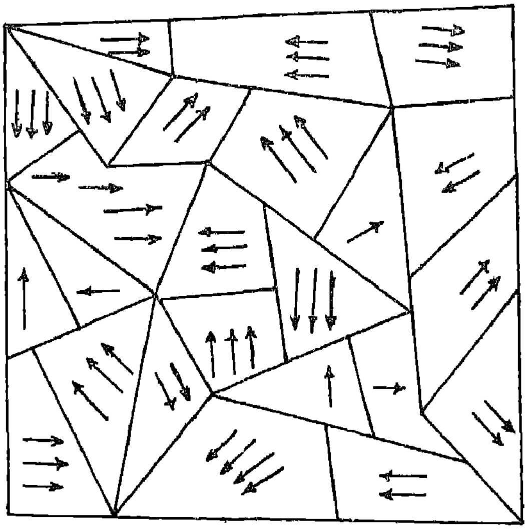

Fig. 17.20 The orientation of domains is such that a piece of ferromagnetic material appears unmagnetised

Exercises

17.1 A 4 m long aluminium wire whose diameter is 3 mm is used to support a mass of 50 kg. What will be the elongation of the wire. Y for aluminium is 7 × 10^{10} N/m^2 (3 96 mm)

17 2 A cable is replaced by another one of the same length and material but of twice the diameter. How will this affect the elongation under a given load ? How does this affect the maximum load it can support without exceeding the elastic limit ?

17 3 A wire increases by 10^{-3} of its length when a stress of 1 × 10^8 N/m^2 is applied to it. What is the Young's modulus of the material of the wire ? (1 × 10^{10} N/m^2)

17.4 Graphite consists of planes of carbon atoms. Between atoms in the planes there are strong interatomic forces. Between the atoms in different planes there are only very weak forces What kind of elastic properties you expect for graphite ?

17.5 Consider a steel bridge 200 m long in a locality where temperature varies from 243°K to 313°K. Find the change in the length of the bridge for the above seasonal variation in the temperature ?
Coefficient of linear expansion of steel 11 × 10^{-6}/°K. (Ans : 15 4 cm)

17.6 A circular hole of radius 1cm is drilled in a brass sheet kept at 293°K. What will be the diameter of this hole when the sheet is heated to 393°K.

(Ans : 2 cm)

17.7 We have a hollow sphere and a solid sphere of equal radii and of the same material. They are heated to raise their temperatures by equal amounts. How will the change in their volumes due to volume expansion be related ?

17.8 Outdoors in winter why does a piece of metal feel colder than a piece of wood ?

17 9 What would be the effect of temperature on the dielectric constant of a substance containing molecules with permanent dipole moment ?

17.10 What kind of thermal conductivity and specific heat requirements would you specify for a cooking utensil ?

17 11 Why do electrons in insulators not contribute to heat conductivity ?

17.12 Why should alkali halides be diamagnetic rather than paramagnetic ?

UNIT 18

Nuclear Physics

18.1 Atomic Nucleus

In unit 9 we have studied the physics of the atom. In Bohr model of the hydrogen atom the hydrogen nucleus is identified as a single positively charged particle called proton. It has a mass of 1.6726 × 10^{-27} kg. which is about 1836 times the mass of an electron. Its charge is the same as the charge on an electron, viz., 1.60 × 10^{-19} C but is positive. The mass of the total number of electrons in any atom is negligibly small compared with the mass of its nucleus. Therefore, the main contribution to the atomic mass is from the nucleus only. From the results of Rutherford's scattering experiments (Section 9.3) nuclear size is estimated to be of the order of 10^{-14}m whereas the diameter of an atom is of the order of 10^{-10}m. Hence most of the atom is empty. In 1932 James Chadwick discovered neutral particles called neutrons. The mass of the neutron is 1.6749 × 10^{-27} kg which is slightly greater than that of the proton and it has no electrical charge. For instance, oxygen nucleus having a charge of + 8e and mass 16 must contain 8 protons. The rest of the mass is accounted for by assuming the presence of 8 neutron's in the nucleus. The constitution of nuclei more complex than hydrogen became clear after the discovery of the neutron. Helium nucleus has a charge + 2e and mass 4 units. Therefore helium nuclei contain two protons and two neutrons each. Like wise Lithium nucleus which has charge + 3e and mass 7 units contains three protons and four neutrons.

18.2 Isotopes

Protons and neutrons which are present in the nuclei of atoms are referred to collectively as nucleons. The number of protons in a nucleus is called its atomic number 7. The total number of nucleus in the nucleus is generally denoted by A. Hence A — Z gives the number N of neutrons in a nucleus. The chemical properties of elements are determined entirely by the electronic configuration of their atoms. We often come across atoms whose electronic structures are identical but whose nuclear masses differ. They are called isotopes. For example, Deuterium nucleus (Deuteron) has one proton and one neutron and Tritium nucleus (Triton) contains one proton and two neutrons. Since Deuterium and Tritium have identical electronic structure like hydrogen they are called *Isotopes of hydrogen.*

Since the neutron has nearly the same mass as the proton, the number A, which

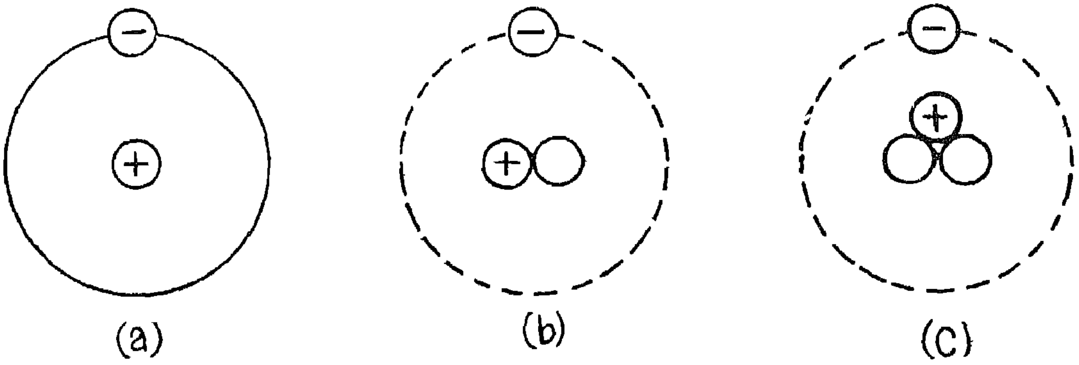

Fig. 18.1 Isotopes of Hydrogen.

represents the sum total of the protons and neutrons in the nucleus is referred to as the mass number of the isotope. The conventional way of representing any isotope is by $\overset{A}{\underset{Z}{x}}$ where x is the chemical symbol of the element, Z its atomic number and A the mass number of the isotope. According to this notation hydrogen is designated as ${}_{1}H^{1}$ and its isotopes deuterium and tritium ${}_{1}H^{2}$ and ${}_{1}H^{3}$ respectively (Fig. 18.1) Lithium has two stable isotopes ${}_{3}Li^{6}$ and ${}_{3}Li^{7}$ (Fig 18.2) and the isotopes of chlorine are ${}_{17}Cl^{35}$ and ${}_{17}Cl^{37}$. Normal chlorine is a mixture of these two isotopes and has an atomic weight 35 45.

When an atom is talked of with particular reference to its nuclear composition it is called a *nuclide*. So istopes are nuclides having the same value of Z. Those having the same value of the atomic mass number A but different values of Z are called *Isobars.*

18.3 Mass Spectrometer

A mass spectrometer is an instrument for measuring the atomic masses. Fig. 18 3

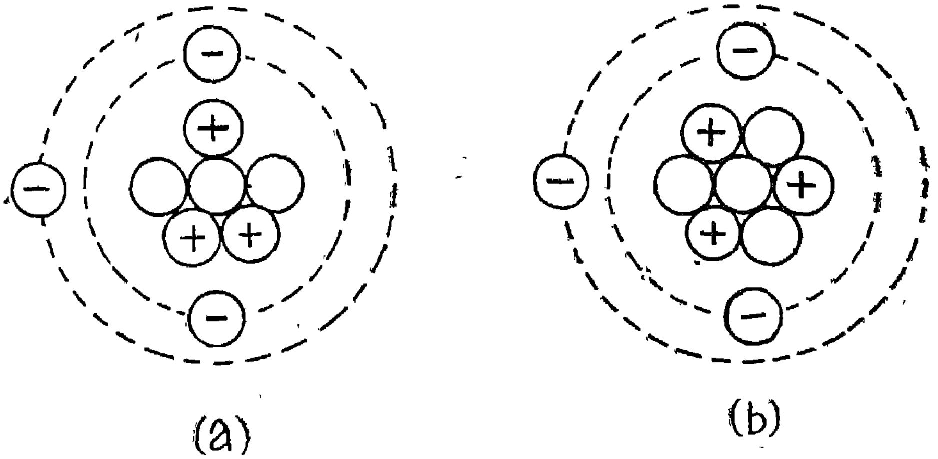

Fig. 18.2 Isotopes of *Lithium*

indicates the arrangement in one type of mass spectrometer. A beam of positive ions from a source is allowed to pass through a long tube T with slits S_1 and S_2 at each end. As they enter through slit S_1 into the tube T, the ions are subjected to electric field E. These ions are also simultaneously subjected to a magnetic field B^1 which is at right angles to the electric field. The sense of the electric field is such that the ions are deflected from left to right whereas the sense of the magnetic field B^1 (into the plane of paper) is such that it deflects the particles from right to left. Thus the ions experience two opposing forces of magnitude Ee and B′eV. As the slit S_2 is exactly opposite to S_1 only those undeflected ions will pass through S_2 into the semi circular chamber of the spectrograph. This condition is satisfied only by those ions having a velocity V which satisfies the condition:

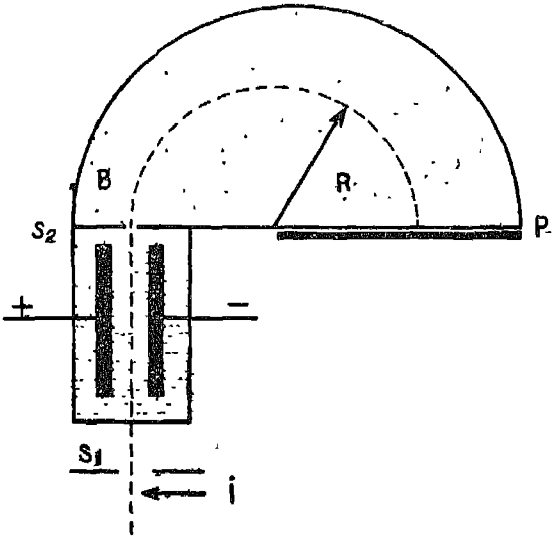

Fig. 18.3 A mass spectrometer.

$$Be'V = Ee \text{ or } V = \frac{E}{B'} \qquad \ldots (18.1)$$

In the region corresponding to the semi-circle in Fig. 18 3 there is only uniform magnetic field B perpendicular to the path of the charged particle. Hence it moves in a circular path and the magnetic force on the particle provides the necessary centripetal force. The radius R of the path of the particle is given by

$$\frac{mV^2}{R} = BeV$$

$$\text{and } R = \frac{m}{e}\ \frac{V}{B} \qquad \ldots (18.2)$$

Substituting the value of V from 18.1 in 18.2 we get

$$\frac{e}{m} = \frac{E}{B\,B^1} \cdot \frac{1}{R} \qquad \ldots (18.3)$$

R is determined by keeping a photographic plate as shown in Fig. 18.3. From a knowledge of the values of E, B and B^1 and measuring R, the value of $\frac{e}{m}$ can be obtained from equation (18.3). Using the known value of unit charge e the atomic mass m can be calculated.

The discovery of isotopes has cleared the apparent anomaly of fractional atomic mass numbers.

18.4 Nuclear Mass

From scattering experiments the radii of the nuclei of oxygen and lead were found to be 3×10^{-15}m and 7×10^{-15}m respectively. The masses of these nuclei are 2.7×10^{-26} kg. and 3.4×10^{-25}kg Let us calculate the densities of their nuclei.

The volume of the nucleus can be found making use of the formula $\frac{4}{3}\pi r^3$ for volume of a sphere of radius r. And density is given by $\frac{\text{mass}}{\text{volume}}$.

Density of oxygen nucleus

$$= \frac{3 \times 2.7 \times 10^{-26}}{4\pi (3 \times 10^{-15})^3}$$

$$= 2.4 \times 10^{17} \text{kg/m}^3$$

Density of lead nucleus

$$= \frac{3 \times 3\,4 \times 10^{-25}}{4\pi (7 \times 10^{-15})^3}$$

$$= 2\,4 \times 10^{17} \text{kg/m}^3$$

We find that the density is the same for both nuclei. The value of this density is very high.

According to the liquid drop model of an atomic nucleus different nuclei are considered as minute droplets of nuclear matter. The fact that the density is the same in the case of both the nuclei in the above example is in accordance with the nuclear theory based on *liquid drop model.* The mass of one cubic centimeter of nuclear matter packed together will be 240 million metric tons.

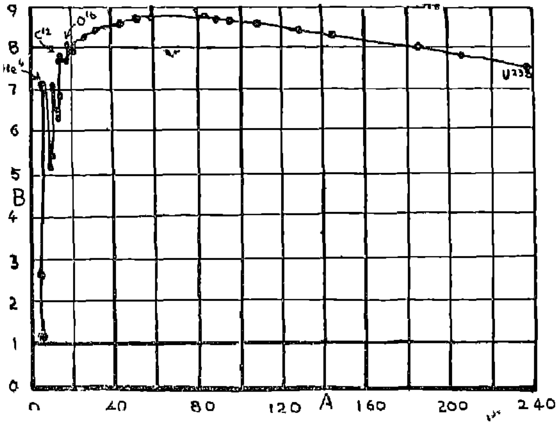

Fig. 14.4 Binding energy per nucleon plotted as a function of mass number A.

18.5 Electron—Volt and Atomic Mass Unit

An electron—volt (ev) is the amount of energy that would be imparted to an electron (charge $1\,602 \times 10^{-19}$C) on being accelerated through a potential difference of 1 volt.

Hence $1 \text{ ev} = 1\,602 \times 10^{-19} \times 1$ J

$= 1\,602 \times 10^{-19}$ J

The size of this unit is convenient for measuring energies in atomic reactions. For example, to remove the one electron from an atom of hydrogen 13 5 ev of energy is needed.

Nuclear forces are very much larger than atomic forces Therefore, while dealing with nuclei, a much bigger unit, a million — electron—volt is used. It is denoted by Mev and is equal to 10^6 ev. For example, about 2.2 Mev energy will be needed to separate the neutron from the proton in the nucleus of heavy hydrogen (Deuterium)

At present atomic masses are expressed in terms of the actual mass of $_6C^{12}$ isotope of carbon. The unit of atomic mass or the atomic mass unit (abbreviated as amu) is $\frac{1}{12}$ of the actual mass of the $_6C^{12}$ isotope of carbon. This is equal to $1\,66 \times 10^{-27}$kg which is close to the mass of an atom of hydrogen. We can now calculate the energy equivalent of one amu using Einstein's mass-energy equation

$$E_0 = m_0 c^2$$

Where E_0 is the energy equivalent to a rest mass m_0 and c is the speed of light. From the above equation

$1 \text{ amu} = (1.66 \times 10^{-27})\ (2.998 \times 10^8)^2$ J

on dividing this by 1.602×10^{-19} J, we get energy in electron volts as equal to 931×10^6 ev or 931 Mev.

$\therefore$ 1 amu = 931 Mev.

18.6 Binding Energy

Deuteron is the deuterium nucleus and it consists of only one proton and one neutron.

Neutron Mass = 1.008665 amu
Hydrogen Mass = 1.007825 amu
Total Mass = 2.016490 amu

But the atomic mass of deuterium is found to be 2.014103 amu which is less than the mass of the particles put together. The difference is 0.002387 amu. In energy units it is equal to 2.22 Mev. This is accounted for as the energy that binds the two nucleons together. The surplus energy which nucleons give up by virtue of their mutual attractions when they become bound together to form a nucleus is called the *binding energy.* When a neutron and proton come together a small part of their mass is radiated in the form of electro-magnetic radiation. The two nucleons attract each other so strongly that, it has been found experimentally, that only γ rays of energy 2.22 Mev or greater can pull the proton and neutron apart.

In the above example of deuteron the binding energy is 2.22 Mev. The total binding energy of Neon $_{10}Ne^{20}$ nucleus is 160.6 Mev and that of Bismuth $_{83}Bi^{209}$ nucleus is 1640 Mev. If these binding energies are divided by their respective mass numbers 20 and 209 we get 8.0 Mev and 7.8 Mev respectively. They represent the binding energy per nucleon in each case. Even though the two nuclides that are considered lie almost at the extremes of the mass number scale their binding energies per nucleon are approximately equal. In Fig. 18.4 the binding energy per nucleon is plotted against mass number. With the exception of $_2He^4$, $_6C^{12}$ and $_8O^{16}$ the values lie on or in the proximity of a curve. The curve has a broad maximum close to the value 8.5 Mev/nucleon in the mass number range from about 40 to 120. As the mass number increases the binding energy per nucleon gradually decreases and falls to 7.6 Mev/nucleon for uranium. The decrease in binding energy per nucleon for large mass numbers is due to the coulomb repulsion between the protons which makes the nuclei increasingly less stable.

18.7 Nuclear Forces

What is the type of force that is responsible for the binding of nucleons in a nucleus? The infinitely small size of the nucleus in which all the protons and neutroes are closely packed indicates that there must be strong attractive forces to hold nucleons (protons and neutrons) together. These attractive forces cannot be electrostatic forces, for protons, being of like charge would havg repelled when held close together and neutrons have no charge. The electrostatic forces between protons being repulsive or disruptive, they cannot account for the strong attractive forces between nucleons which are responsible for the large binding energy in a nucleus of the order of 8 Mev per nucleon. The gravitational force is a force of attraction between every pair of nucleons but it is infinitesimally smaller than the powerful attractive forces between nucleons. It is, therefore, assumed that the purely attractive forces which exist between nucleons are of a new type which have no analogy with other forces such as gravitational and electrostatic forces in classical physics. These attractive forces are called *nuclear forces.* We shall now consider some important characteristics of nuclear forces. The electrostatic coulomb forces operate between any pair of charges at any distance following inverse square law. The situation in the case of the nucleus is entirely different. All the nucleons are closely packed in the tiny nucleus like a set of marbles in a box. The forces which hold the nucleons together must be those which exist between the individual neighbouring nucleons in the nucleus which are separated by distances of the order of 10^{-15}m. The nuclear forces between nucleons are, therefore, *short range forces* operating over very short distances only of the order of 10^{-15}m.

These short range forces operate only between two neighbouring nucleons and they are of three types. These are — (1) the force between a proton and neutron (pn force) (2) the force between two protons (pp force), and (3) the force between two neutrons (nn force). All these are attractive forces. It may be noted that, in the case of (pp) nuclear force, there is also a repulsive force between the two protons, but this is weak compared to the strong nuclear force. There is enough experimental evidence to show that the (pn), (pp) and (nn) nuclear forces are approximately equal. This is the *charge independence character* of nuclear forces.

18.8 Nuclear Reactions

An important landmark in the progress of nuclear physics was the discovery of artificial transmutation of elements by Rutherford in 1919. Transmutation is transformation of one element into another. He was the first to artificially transmute or change the element nitrogen into oxygen by bombarding nitrogen nuclei with high energy α particles from a radioactive source. Fig. 18 5 shows the schematic diagram of the apparatus used by Rutherford. T is a tube filled with nitrogen gas. α particles from a radio-active source R were allowed to bombard the nitrogen nuclei. A is a thin silver disc covering the opening at the end of the tube and S is a fluorescent screen for observing scintillations through a microscope M.

When α particles bombarded nitrogen nuclei, scintillation, were ovserved on the screen S due to nuclear particles which were identified as protons. The nuclear reaction that has taken place may be written symbolically as

$${}_2He^4 + {}_7N^{14} \rightarrow {}_8O^{17} + {}_1H^1 \quad \dots \quad (18.4)$$

${}_8O^{17}$ is an isotope of oxygen. Starting with two stable nuclei namely, helium and nitrogen, two new stable nuclei oxygen and hydrogen were produced. This process is, therefore, referred to as artificial transmutation of elements.

Conservation Laws in Nuclear Reactions

It is found that in any nuclear reaction there is conservation of (a) Momentum (b) Charge, and (c) Energy. (a) In any nuclear reaction the total momentum of the Atoms entering into reaction remains the same as the total momentum of the products after the reaction.

(b) In the nuclear reaction stated above

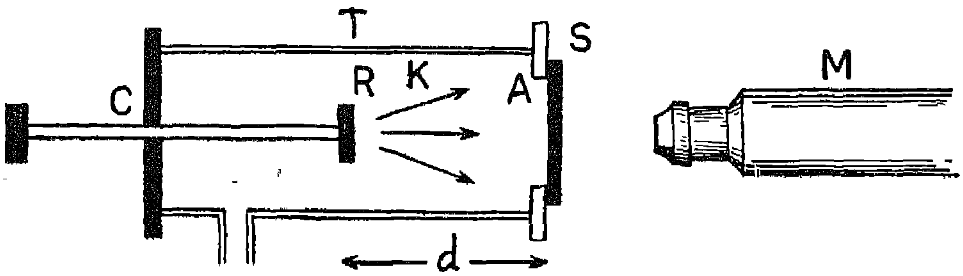

Fig. 18.5 Schematic diagram of Rutherford's apparatus.

the total charge of the items entering into reaction is 2 + 7 = 9 units and that after the reaction is 8 + 1 = 9 units. Therefore, the total charge is conserved

(c) In relativistic mechanics the energy of any particle is the sum total of rest-mass energy and Kinetic energy The combination is effected using Einstein's equation $E_0 = m_0c^2$ which expresses the equivalence of mass and energy According to the principle of conservation of energy the total energy before a reaction equals the total energy after the reaction. Nuclear energy equations are generally written in atomic mass units. But any consistent units of mass or energy can be used

The mass energy equation for the nuclear reaction just studied may be written as :

$m_1c^2 + {}^Ek_1 + m_2c^2 + {}^Ek_2 = m_3c^2 + {}^Ek_3 + m_4c^2 + {}^Ek_4$ ... (18.5) Where m_1, m_2, m_3, m_4, are the rest-masses of H_2e^4, N^{14}. $_8O^{17}$ and $_1H^1$ respectively and ${}^Ek_1$, ${}^Ek_2$, ${}^Ek_3$, and ${}^Ek_4$ their respective Kinetic energies

Note : If, initially, energy is also supplied in the form of a photon and if a photon is emitted finally, then the energies of these photons also should be included in the equation.

Let us express all the quantities in the equation in atomic mass units. The Kinetic energy ${}^Ek_2$ of $_7N^{14}$ is very small (fraction of one Mev). Hence it can be neglected in comparison with other quantities. The energy of the α particle in this experiment was found to be 7.7 Mev. Expressed in atomic mass units.

$$7.7 \text{ Mev.} = \frac{1}{931} \times 7.7 = 0.008270$$

amu neglecting ${}^Ek_2$, the masses of the items on the left hand side of equation (18.5) are:

$m_1 = {}_2He^4 = 4.002604$ amu
$m_2 = {}_7N^{14} = 14.003074$ amu
${}^Ek_1 = 0.008270$ amu
Total $= 18.013948$ amu

Coming to the right side of the equation

$m_3 = {}_8O^{17} = 16.999133$ amu
$m_4 = {}_1H^1 = 1.007825$ amu
$m_3 + m_4 = 18.006958$ amu

Substituting the above values in equation 18.5 we get

$$\begin{aligned} {}^Ek_3 + {}^Ek_4 &= 18.013948 \ldots 18.006958 \\ &= 0.006990 \text{ amu} \\ &= 931 \times 0.006990 = 6.508 \text{ Mev.} \end{aligned}$$

6.508 Mev. energy is shared as kinetic energy by the proton and oxygen nucleus.

Chadwick in 1932 conducted an experiment in which he allowed α particles (D) from polonium to strike beryllium metal (B) (Fig. 18.6). Penetrating rays were found to come out of the beryllium metal. They were found to be particles without charge. They are called neutrons (N). The reaction may be written as

$${}_2He^4 + {}_4Be^9 \longrightarrow {}_0n^1 + {}_6C^{12} \quad \ldots (18.6)$$

$_0n^1$ denotes the neutron since its charge is zero and mass nearly the same as that of a proton.

Instead of α particles high speed protons also can be used as projectiles to study the disintegration of nuclei. In one experiment

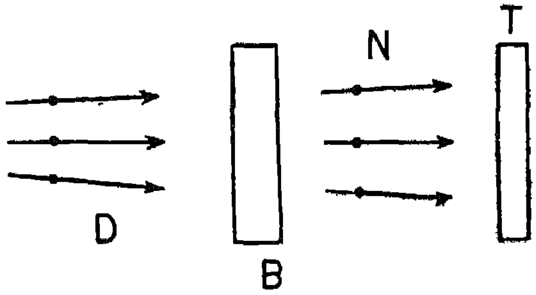

Fig. 18.6 Production of neutrons by allowing alpha particles to strike beryllium metal.

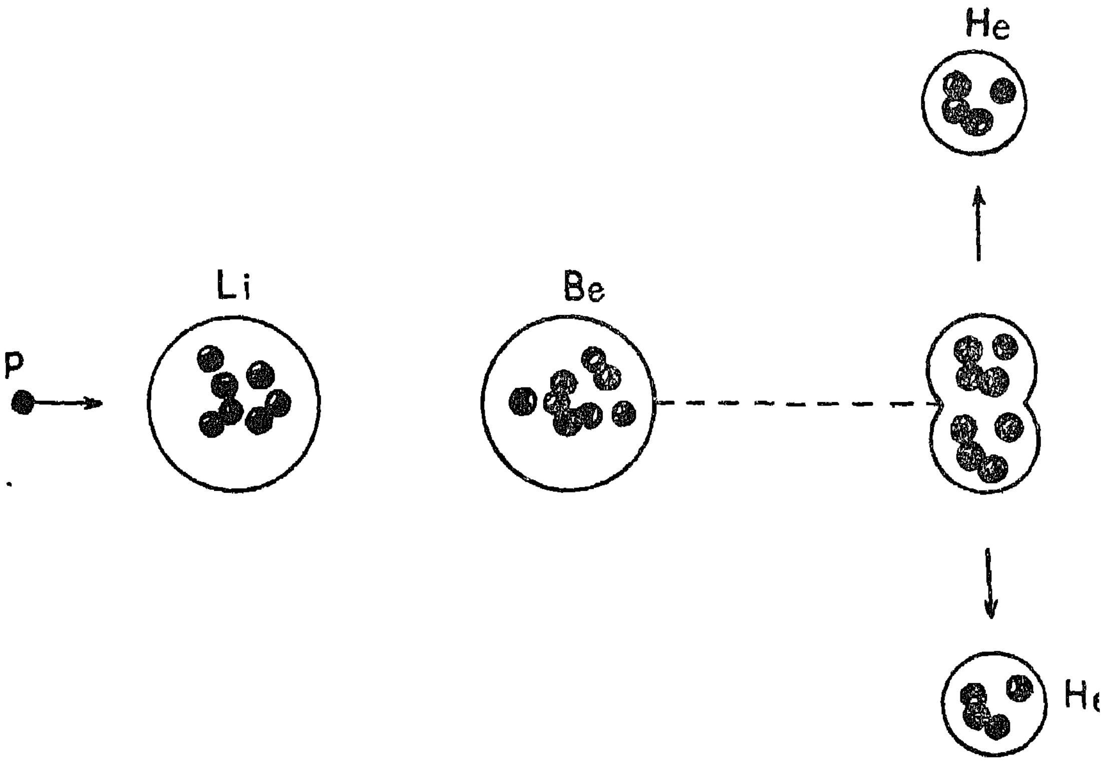

Fig. 18.7 Disintegration of a lithium nucleus by a proton.

protons accelerated to energies of the order of 0.15 Mev were used to bombard Lithium metal. When a proton collides with a lithium nucleus and is captured, two α particles are produced (Fig. 18.7). Neglecting E_{k2} the reaction may be writtens as

$$_1H^1 + {}_3Li^7 + E_{k1} \longrightarrow 2\,({}_2He^4) + 17 \text{ Mev.} \quad \dots (18.7)$$

The energy of each α particle was measured to be about 8.5 Mev.

18.9 Properties of Neutrons

We have seen that a neutron is a particle with no charge and with a mass slightly higher than that of the proton. Neutrons are present in the nuclei of all elements except hydrogen. Although in a nucleus they appear to last for ever neutrons outside a nucleus have an average life of about 1000 seconds before they decay. A free neutron spontaneously decays into a proton. electron and an antineutrino.*

$${}_0n^1 \longrightarrow {}_1H^1 + {}_{-1}e^0 + \upsilon^{-0}$$

The electron charge being —1 the subscript to e is —1. The mass of an electron being about $\frac{1}{1836}$ proton, it is negligibly small compared to the mass of a proton. Hence the mass of electron is written as O.

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* antineutrino is an elementary particle (See-Unit. 19)

Being without charge neutrons are not subject to Coulomb forces and can easily penetrate even heavy nuclei. Therefore, neutrons serve as ideal projectiles for inducing nuclear reactions in the study of the properties of nuclei. In the nuclei of heavier elements there are more neutrons than protons for the following reason. Protons being positively charged repel each other and in order to maintain stability of the nucleus more neutrons become necessary in the case of heavier elements

By keeping a thin plate of beryllium metal in the path of an intense beam of deuterons from a cyclotron (an instrument used for accelerating atomic particles) high speed neutrons are produced. Substances of known chemical composition are inserted in the path of these neutrons and the disintegration products are studied.

Following are some examples of neutron reaction using high-speed neutrons :

$$_0n^1 + {}_9F^{19} \longrightarrow {}_7N^{16} + {}_2He^4$$
$$_0n^1 + {}_{27}Co^{59} \longrightarrow {}_{27}Co^{60} + \gamma \text{ ray}$$
$$_0n^1 + {}_{80}Hg^{198} \longrightarrow {}_{79}Au^{198} + {}_1H^1$$

The last reaction indicates the transmutation of mercury into gold. Thus the alchemists' dream of producing gold from mercury has been realised However, this method cannot be adopted, in practice, for obtaining gold from mercury because it would be very expensive.

Thermal Neutrons : Since a neutron does not experience a repulsive force as it approaches a nucleus the chances of its penetrating the nucleus are more. Hence disintegration of a nucleus by slow moving neutrons is possible while low speed charged particles like protons, deuterons or α particles cannot easily bring about disintegration of nuclei.

Any material used to slow down fast moving neutrons is called "moderator". Paraffin or any such material having large number of hydrogen atoms serves as a moderator. Deuterium is another example. When two particles of the same mass have an elastic head on collision with one another their velocities interchange Neutrons and protons have nearly the same mass. When fast moving neutrons are passed through paraffin the neutrons make elastic collisions with hydrogen nuclei which have comparatively much smaller velocities and are slowed down. Ultimately the velocities of the neutrons correspond to the velocities of random motion of the atoms and molecules of the moderator. Such neutrons are called *Thermal Neutrons.* We will study about the use of thermal neutrons to produce fission of nuclei of heavier elements like uranium in unit 19

18.10 Artificial Radio activity

The phenomenon of artificial or induced radio activity was discovered by F. Joliot and Irene Curie Joliot. They exposed various substances to α rays from naturally radioactive elements and made extensive study of the disintegration produced in them.

In one experiment α rays from polonium were allowed to strike an aluminium plate (Fig. 18.8) (a) At the end of the experiment, even after the radio-active source was taken away the radiation detector was found to respond to some kind of penetrating radiation They investigated the nature of these radiations and found them to be positrons* coming from aluminium which

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* Positron is a particle having the same mass as an electron but with equal positive charge (See Unit. 19)

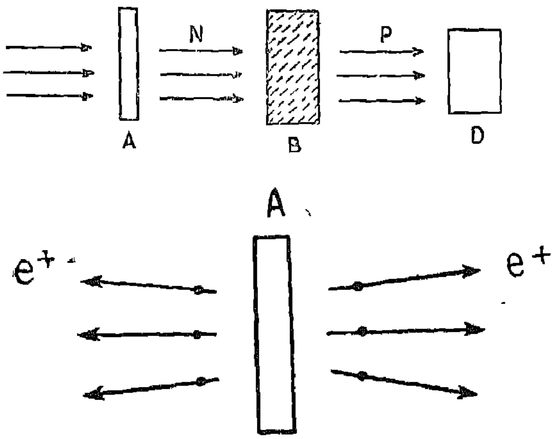

Fig. 18.8 Schematic diagram of the experimental arrangement used by the Curie-Joliots in their discovery of induced radioactivity. N—Neutrons, A—Aluminium, P—Protons, B—Paraffin block, D—Detector

was formerly exposed to α rays (Fig. 18.8). Further, the emission of positrons decreased with time according to an exponential law. Hence the phenomenon was just like natural radio-activity.

When α rays strike aluminium, neutrons are ejected out and phosphorus $_{15}P^{30}$ is formed according to the reaction

$$_2He^4 + {}_{13}Al^{27} \longrightarrow {}_{15}P^{30} + {}_0n^1 \quad ...(18.8)$$

It is this isotope $_{15}P^{30}$ of phosphorus which exhbited radio-activity Its half-life is 2.55 minutes. It decays into silicon emitting a positron accompanied by nutrino**

$$_{15}P^{30} \longrightarrow {}_{14}Si^{30} + {}_1e^0 + \upsilon$$

18.11 Radio-Isotopes

In the previous section we have seen that radio-active phosphorus $_{15}P^{30}$ was obtained when aluminium was bombarded by α particles. A variety of radio-active isotopes are produced by adopting similar procedure using besides α particles, deuterons, neutrons, etc,. as projectiles to hit several elements as targets.

Although produced in very small quantity, it is possible to separate the radio-active phosphorus from the aluminium (reaction studies in 18.10) which is non-radio-active. Also it is possible to trace this small amount of radio-active phosphorus through successive chemical processes by using suitable detectors. This fact is made use of in radioactive "tracer techniques". In nuclear reactors, described in unit 19, large number of radio-active isotopes can be produced. Some radio-isotopes have great application as tracers in various fields such as medicine, agriculture and industry. Besides their use as tracers, radio-isotopes have a variety of other applications. In the next section some examples of the use of radio-active isotopes are given.

18.12 Use of Radio-isotones

(*a*) *In Medicine*

In medical diagnosis radio isotopes are particularly useful in cases where conventional diagnostic procedures are not quite helpful.

When Iodine is taken into the body, it is absorbed more strongly by the thyroid gland than by other tissues. The radio isotope $_{53}I^{131}$ has a half-life of 8 days and it decays ley β-emission. Radioiodine is administered to a patient and by means of sensitive detectors of β rays, information about the size, activity and functioning of the thyroid gland is obtained.

**Neutrino is an elementary particle which carries no charge and has no rest-mass (See unit. 19)

Diagnosis of brain tumors is difficult because of the inaccessibility. Radioiodine and some other isotopes are found useful in the diagnosis of brain tumors. A small quantity of the radioisotope is incorporated in organic dyes which are absorbed strongly by the tumor tissue than by normal tissues. This is injected into the body Then by means of detectors of the emitted radiation the radiologist gets information about the size and location of the tumor.

Suppose the circulation of blood in a person is restricted. It is easy to locate the position of the constriction by using Radiosodium. Sodium chloride prepared by using radiosodium is taken and a solution of it of proper strength is injected into a vein of the patient. A radiation counter is kept in contact with, say, one of the feet of the patient. If the blood circulation is normal the radio-activity increases rapidly and reaches a maximum value in less than one hour. If the radioactivity increases slowly it means there is difficulty for the blood to reach the foot. By shifting the detector to different parts of the body the exact spot where the constriction exists can be located and necessary treatment provided.

Radiation has long been used in medical therapy for controlling the development and growth of cells, for example, in the treatment of some forms of cancer In certain cases small capsules or needles containing radium (or its decay products) are implanted within the diseased organ which is thus subjected to the action of gamma rays At the present time artificial radioisotopes provide more convenient and cheaper alternative sources of radiation both in teletherapy and for implanting in the body The penetrating rays from the radio-isotope of cobalt Co^{60} have been found to permit higher doses to tissue deep in the body with a minimum of skin reaction.

Leukemia disease is found to respond to treatment by radiation from radio-isotopes of phosphorus. Radiation from radioisotope of iodine has been found to be effective in the treatment of *hyperthyroidism.* It may be noted here that radioisotopes have been used for internal irradiation by taking advantage of the preferred absorption of certain elements in particular organs or tisssues of the body.

(*b*) *In Agriculture*

Phosphorus is applied to soil to improve plant growth. Various types of phosphate fertilizers are available and it is helpful to know which variety is utilized most effectively by the particular soil. By using a fertilizer in which radiophosphorus is employed it is possible to know to what extent the plant absorbs phosphorus from that particular fertilizer. This facilitates the selection of the fertilizer most suitable to the soil. Without the use of radio phosphorus it is not possible to distinguish between the phosphorus from the soil and that from the fertilizer.

Crop Mutations : Plant genetic studies have helped in obtaining crops with high yield, resistance to disease and adaptability to new environment Desirable mutations are induced by irradiation of seeds by γ rays or neutrons. High quality wheat and rice mutants resistant to certain diseases and several jute mutants have been evolved by this process.

Certain insect pests can be controlled by adopting the so-called sterile male technique. The procedure consists of laboratory breeding of large numbers of male insects, sterilizing them with α radiation and releasing them for mating in the infested area This will bring about rapid reduction in the population of the insect. In the U.S A. the screw worm fly, a pest which caused great damage to livestock, was eliminated by adopting the above procedure.

(c) *In Industry*

Radioactive tracers have wide application in industry both in research and process control. When petroleum products are transported through a single long pipeline, radioactive Antimony-124 used as a tracer, is injected at the source between two different oil stocks and a detector kept outside the pipeline at the receiving end of the line announces the arrival of the new stock of oil. In a similar way the velocity of flow and the degree of intermixing of oil stocks can also be found out

The purpose of lubricants is to minimise the wear of surfaces, for example of gears and bearings in machinery. It is necessary to select a suitable lubricant for this purpose. A small quantity of a radioactive isotope of iron may be incorporated in the steel of a bearing under test. The lubricating oil is drawn and checked periodically for radioactivity caused by the wearing away of the steel bearing. This gives and idea of the effectiveness of the lubricant.

Gamma rays from a Co^{60} source are found to have the effect of a catalyst in the combination of hydrogen bromide and ethylene Large quantities of ethyl bromide are produced by this process. Polyethylene, which has great commercial importance, is also produced by Cobalt-60 gamna ray bombardment of ethylene. Irradiated polyethylene has high heat and electrical resistance properties and serves as excellent electrical insulating coating for wires.

In industry quality control of castings and inspection of welded pipes and fabricated machine parts can be made by radiography using cameras housing radio-isotopes like iridium-192 and cobalt-60.

Radio-tracer techniques have been developed for the study of movement of silt on the sea bed. These studies are of vital importance especially in planning dredging operations near ports. A major study of this kind undertaken in recent years was at the Sethusamudram canal project in the Palk Straits. In hydrology, radio-tracers have been used for studying seepage in dams.

(d) *Radiocarbon Dating*

Carbon dating is a recently developed technique for the dating of cultural history, especially for the dating of items of animal or vegetable origin. C^{14} is a radio-isotope of carbon for which there is "radioactive equilibrium" in the atmosphere. The equilibrium is between production of C^{14} by cosmic rays and its diminution by radioactive disintegration. The half-life of this isotope is 5600 years. All animals and plants have a certain amount of C^{14} which remains in equilibrium with the C^{14} in the atmosphere. After the death of an organism the C^{14} in the body keeps on disintegrating as per exponential decay formula without further replenishment of C^{14}. For example, old bones contain less C^{14} than new bones and old wood contains less C^{14} than new wood. By measuring the ratio of concentration of C^{14} to C^{12} in the remnants of the dead organism and comparing it with the normal ratio in a living organism, it is possible to estimate how long ago the organism ceased to live. This method is now widely used by anthropologists and archaeologists and even historians for finding the age of ancient excavations, manuscripts etc. This method which has become a standard technique is known as radioactive dating.

Exercises

(Where necessary the following data may be used :
Mass of $_0n^1$ = 1.008 665 amu; mass of $_1H^1$ = 1.007 825 amu ;
mass of $_2He^4$ = 4 002 604 amu)

18.1 What is 'atomic number' ? What are 'isotopes' ?

18 2 What is the number of protons and the number of neutrons in each of the following nuclei :
$_3Li^6$; $_5B^{10}$; $_6C^{12}$; $_{10}Ne^{22}$; $_{15}P^{31}$; $_{38}Sr^{88}$;
$_{56}Ba^{137}$; $_{92}U^{238}$

18 3 The mass of $_{17}Cl^{35}$ is 34 9800 amu. Calculate its binding energy. What is the binding energy per nucleon ?

(288 Mev ; 8 2 Mev)

18 4 Calculate the binding energy per nucleon in the nuclei $_2He^4$ and $_{26}Fe^{56}$ given mass of $_{26}Fe^{56}$ = 55.934 932 amu.

(7 1 Mev ; 8 8 Mev)

18 5 Calculate the binding energy per nucleon for $_{15}P^{31}$ given mass of $_{15}P^{31}$ = 30 973 763 amu

(8.5 Mev)

18 6 Calculate the energy released in the reaction :

$$_3Li^6 + {}_0n^1 \rightarrow {}_2He^4 + {}_1H^3$$

given mass of $_3Li^6$ = 6 015 126 amu and that of $_1H^3$ = 3.016 049 amu

(4.78 Mev)

18 7 A neutron strikes a $_5B^{10}$ nucleus with the subsequent emission of an alpha particle. What is the atomic number, mass number and chemical name of the remaining nucleus ?

18.8 Describe briefly Rutherford's experiment in which he discovered artificial transmutation of elements ?

18.9 When neutrons bombard $_{16}S^{32}$ nuclei, radioactive phosphorus $_{15}P^{32}$ is produced. Write down the reaction equation.

18.10 When $_{14}Si^{28}$ nuclei are bombarded by neutrons, protons issue out from the target. Write down the reaction equation.

18 11 Complete the following nuclear reaction equations :

(i) $_{13}Al^{27} + {}_0n^1 \rightarrow {}_2He^4 +$
(ii) $_{15}P^{31} + \gamma \rightarrow {}_0n^1 +$
(iii) $_9F^{19} + {}_1p^1 \rightarrow {}_8O^{16} +$
(iv) $_6C^{14} \rightarrow {}_{-1}B^0 +$

18.12 What are 'thermal neutrons' ? Explain the action of a moderator in slowing down fast moving neutrons.

18.13 Write a note on the use of radio-isotopes in
(a) medicine, (b) agriculture and (c) Industry.

Nuclear Interactions

19.1 Fission

In 1939 Otto Hahn and his collaborators discovered that when a uranium nucleus, bombarded with neutrons, captures a neutron it splits into two nearly equal fragments. Since this process resembles fission of cells in biology, the phenomenon is called nuclear fission.

Fig. 19.1 shows the splitting of a

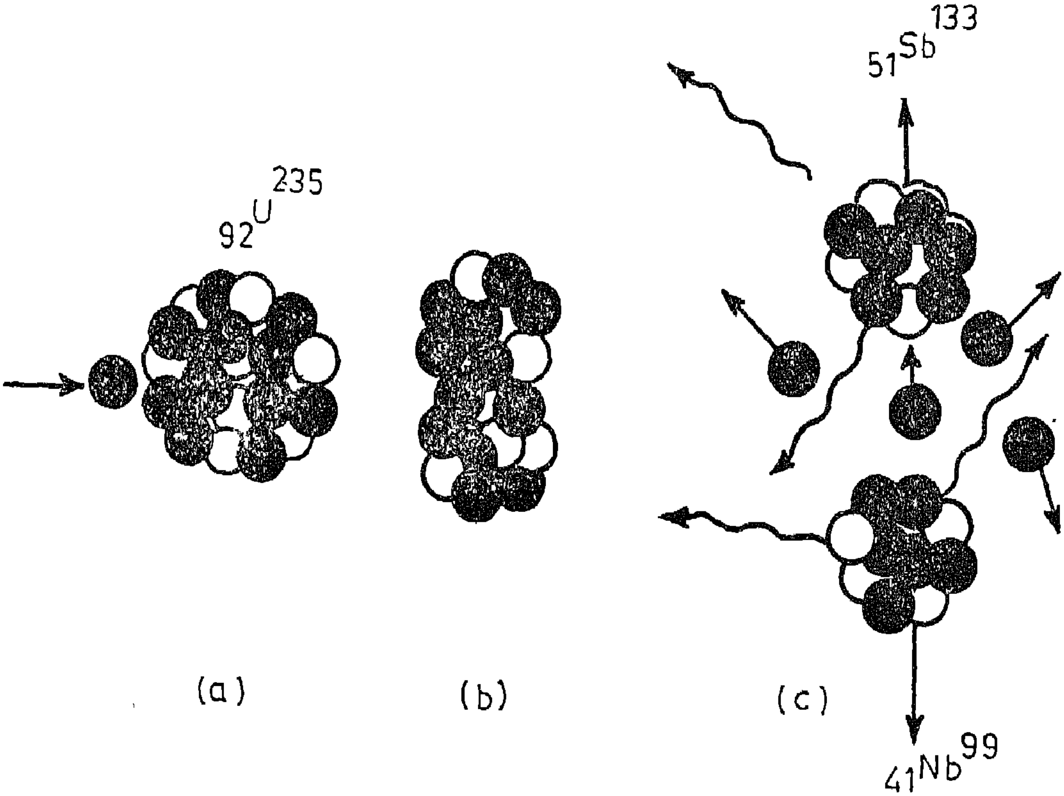

Fig. 19.1 Fission of a nucleus of Uranium $_{92}U^{235}$

nucleus of ${}_{92}U^{235}$ after it is bombarded by a neutron. The resulting fragments were identified as the nuclides of ${}_{51}Sb^{133}$ and ${}_{41}Nb^{99}$. Both these nuclides are unstable and by radioactive disintegration reduce to the stable nuclides ${}_{55}Cs^{133}$ and ${}_{44}Ru^{99}$ respectively. Fission is accompanied by the release of neutrons and α rays.

Uranium found in nature has three principal isotopes. They are as follows:

Isotope	Relative abundance (Per cent)
${}_{92}U^{238}$	99.280
${}_{92}U^{235}$	0.714
${}_{92}U^{234}$	0.006

All the three nuclides are radioactive and they decay by alpha emission. Energy liberation during the fission of uranium is due largely to the isotope U^{235} when slow neutrons are used.

19.2 Energy released in fission

The fission fragments and the accompanying neutrons are released with high velocities. So, large kinetic energies are associated with them. Also energy is released in the form of rays. Let us make an estimate of the average energy liberated in the fission of U^{235} nucleus. Let the two fragments have mass numbers say, 100 and 133; then the number of neutrons released is three.

Before fission :

Mass of U^{235}	235 043933 amu
Mass of one neutron ...	1.008665 amu
Total	236.052598 amu

After fission :

Total mass of the two fragments	232.812000 amu
Mass of 3 neutrons ...	3.025995 ,,
Total	235 837995 ,,

Hence the mass converted into energy is 0.214603 amu, its equivalent in energy units is 200 Mev. nearly. This energy is accounted for as the energy of the gamma rays and the kinetic energy of the fission fragments and the released neutrons.

19.3 Chain Reaction and Controlled Chain Reaction

If a neutron bombards a ${}_{92}U^{235}$ nucleus and produces fission it will set off a chain reaction in a lump of ${}_{92}U^{235}$ as follows The nucleus which captures a neutron splits apart with great violence releasing one or more additional neutrons. These neutrons in turn would be captured by the nuclei of nearby uranium atoms and they, in turn, would split, liberating fresh neutrons which again would be absorbed by uranium nuclei resulting in further fission reactions Because of the chain reaction, in a very short time, the number of nuclei undergoing fission multiplies in geometric progression. Fig. 19 2 shows a schematic diagram of a chain reaction. It is found that on an average about 10^{-8} second elapses between the emission of a neutron and its subsequent capture by a uranium nucleus. Assuming that the chain reaction is uncontrolled, that is, if all the neutrons that are liberated are allowed to induce further fission, then in about one microsecond (one-millionth of a second), there will be a release of energy of the order of 2×10^{13} Joules, which is a very large quantity. That is why an uncon

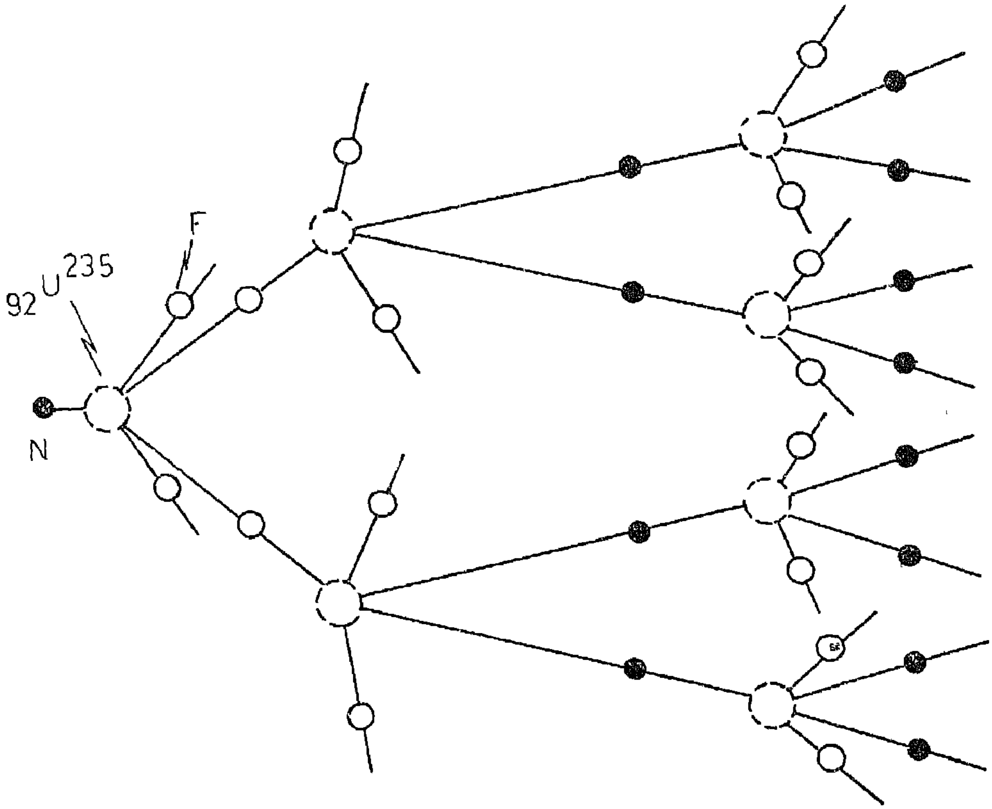

Fig. 19.2 Schematic diagram of a chain reaction in $_{92}U^{235}$.

trolled chain reaction can cause a violent explosion. For this to occur the lump of uranium has to be greater than a certain minimum size

If the mass of uranium is too small then the liberated neutrons have large scope to escape from the surface and the growth of chain reaction may end before it becomes large enough for an explosion Hence, there is possibility for an explosion to take place by chain reaction, only if the mass of uranium is greater than some critical size.

In the atomic bomb that was first tested in the U.S.A., $_{92}U^{235}$ was used as the active substance. The requisite quantity of this substance was taken and it was divided into two or more units, each smaller than the critical size. They were kept separated by a short distance. By means of some mechanism they were brought together at the appointed time. Simultaneously neutrons from some source were made to enter the mass and start the chain reaction, finally resulting in a tremendous explosion and release of enormous quantity of energy.

Whether a mass of active material will sustain a chain reaction or not is determined by what is called the *reproduction factors.* If the rate of neutron production is equal to the rate at which neutrons disappear, the reproduction factor is one and the mass is said to be *critical*

By the use of moderators the fast neutrons, produced during fission, can be slowed down. And by the use of materials like cadmium, which absorb neutrons without disintegration, the chain reaction can be controlled. Thus, by properly controlling the chain reaction, a constant level of power output can be maintained. Such a controlled chain reaction will be a very efficient source of power. If a conventional power plant needs a consumption of 3 tons of coal per day per 1000 kw, the same output of energy can be produced in a nuclear power plant by using 1g of a suitable isotope as fissionable material per day.

19.4 Nuclear Reactors

A nuclear reactor is a device in which nuclear fission can be maintained as a self-sustained, yet controlled chain reaction. The orignal reactors were known as atomic piles. The first atomic pile was put into operation in Chicago in 1942. By using uranium in the atomic pile, products such as neutrons, radioactive isotopes and heat energy were obtained We have, at present, many kinds of nuclear reactors. The main items in a reactor are :

(1) a quantity of fissionable material,
(2) a moderator for slowing down the neutrons to thermal velocities, and
(3) substances like cadmium for controlling the chain reaction

The neutrons released by the fission of uranium have energies of the order of 2 Mev. By the use of moderators they can be slowed down to thermal energies of 0.0253 ev (corresponding to velocities of 2200 m/s.) Fig 19.3 shows the schematic diagram of a reactor made up of large graphite or carbon blocks used as moderator. There are long cylindrical holes through the blocks for the insertion of fuel elements, control rods, materials to be irradiated and detecting devices. The fuel element is pure uranium metal sealed in aluminum cylinders. Boron or Cadmium is used for control rods, because of their ability to capture slow neutrons. The arrangement of parts in a power reactor shown in Fig. 19 4 helps to understand the working of a reactor.

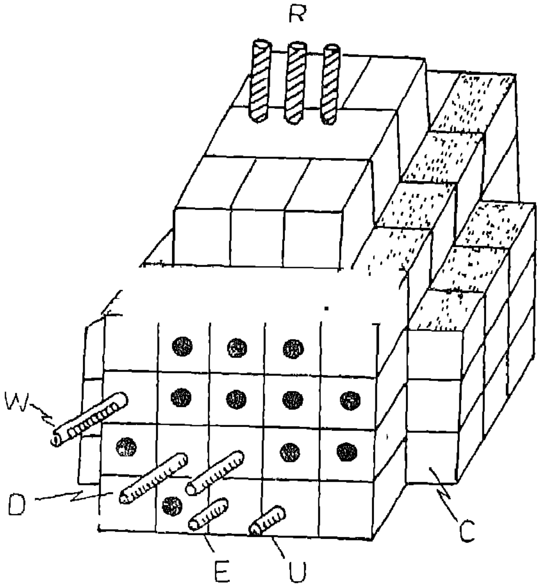

Fig. 19.3 Schematic diagram of an atomic pile (earliest type of nuclear reactor) R—control rods, C—carbon block, W—cooling pipes, D—counting device.

When a few U^{235} nuclei undergo fission fast neutrons are liberated. As they pass through the surrounding carbon blocks they are slowed down. These thermal neutrons are captured by U^{235} nuclei and fission occurs. By proper arrangement of the control rods (by raising and lowering the rods) the fission events are suitably controlled.

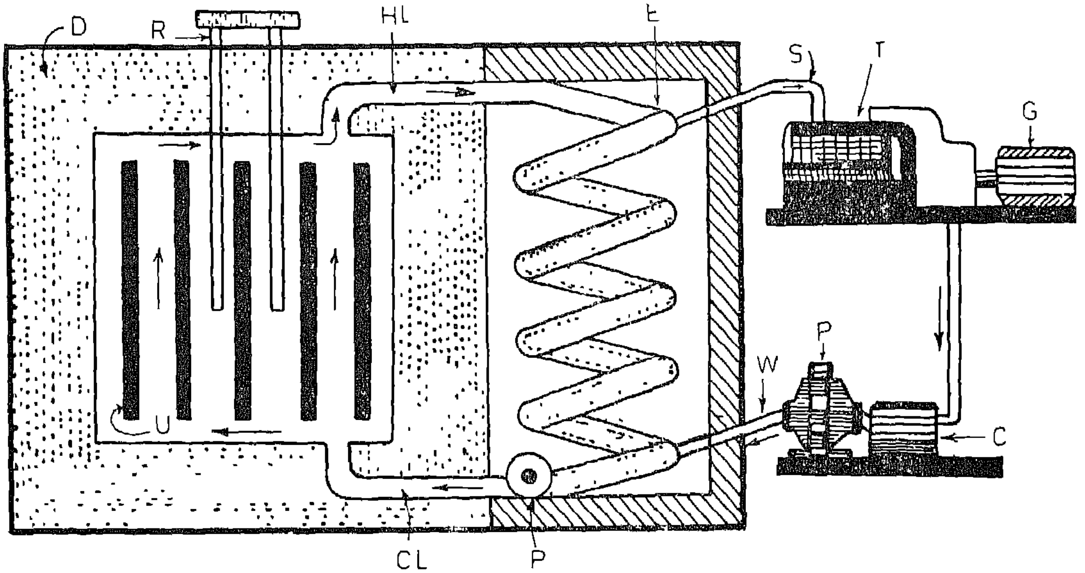

Fig. 19.4 Schematic diagram of a nuclear power plant P—pump, R—control rods, HL—hot liquid, CL—cool liquid, S—steam, C—condenser.

In a nuclear reactor plutonium ($_{94}pu^{239}$) and many other radio-active isotopes can be produced Considerable heat is generated in these reactors. This fact is made use of in the construction of nuclear power plants. The heat generated is used for heating suitable coolants which in turn heat water and produce steam. This steam is made to rotate a steam turbine and thereby drive a generator for production of electric power. Fig. 19 4 shows the schematic diagram of a nuclear power plant.

India's First Reactor Apsara

Fig. 19.5 shows a picture of India's first research reactor called Apsara. This was started in 1956 by the Department of Atomic Energy for a variety of needs in the field of nuclear research as well as for the production of radio-isotopes A large number of radio-isotopes are produced by means of this reactor and supplied to various laboratories and hospitals in India for their use.

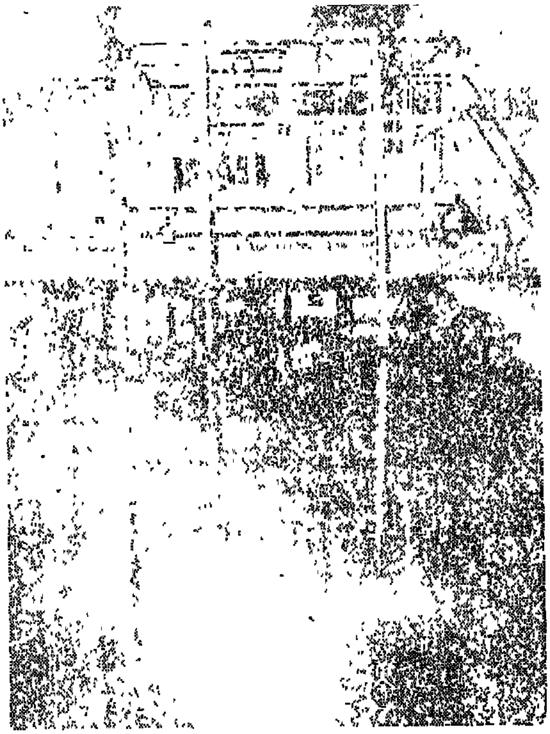

Fig. 19.5 View of Apsara—India's first research reactor at Bhabha Atomic Research Centre, Trombay,

Prior to the commissioning of Apsara in 1956 radiation sources for biological investigations were practically non-existent in India. With the availability of Apsara for research the radiobiological programme in India received a much needed impetus. At the Biology Division of Bhabha Atomic Research Centre, Trombay, neutrons from Apsara are used in genetic studies in agricultural crops. Very useful results have been obtained.

Apsara is a swimming-pool type reactor. A reactor is named according to the moderator or coolant used with it. In the swimming-pool type of reactor a large tank of water is used as a protective shield for the operating personnel. Samples to be irradiated by neutrons are lowered into water and an observer can readily see the blue glow of the water around it (Fig. 19.5). This glow is caused by Cerenkov radiation.*

19.5 Fusion and Energy Generation in Sun and Stars

All animals derive their energy ultimately from plants and plants get their energy from photosynthesis caused by sunlight. Thus all living beings exist because of solar energy. The total radiation emitted by the sun is estimated to be about 3.8 × 10^{26} Joules of energy per second. The sun has been giving out energy, roughly at this rate, for millions of years. Yet there are no signs of cooling. For a long time this was a mystery to the physicists. But nuclear studies have given a clue to the understanding of this problem.

We have seen that it is possible to produce fission of nuclei of heavy atoms like uranium and plutonium. The reverse of fission, namely, fusion of light nuclei, like protons and deuterons, into helium nuclei is also found possible. Like fission processes, already studied, fusion processes also are accompanied by the liberation of large quantities of energy.

The core of the sun is estimated to be at a temperature of about 20 million degrees C. It is presumed, with the knowledge available at present, that, by a set of nuclear reactions (primarily the proton-proton cycle) taking place at this high temperature enormous amount of energy is liberated. The reactions are :

$${}_1H^1 + {}_1H^1 \rightarrow {}_1H^2 + {}_1e^0 + 0.4 \text{ Mev.}$$

$${}_1H^1 + {}_1H^2 \rightarrow {}_2He^3 + 5.5 \text{ Mev.}$$

$${}_2He^3 + {}_2He^3 \rightarrow {}_2He^4 + 2({}_1H^1) + 12.9 \text{ Mev.}$$

Each of the first two reactions has to occur twice to produce one ${}_2He^4$ or helium nucleus. The net result is the fusion of four hydrogen nuclei to form one nucleus of helium atom besides producing two positrons and liberating 24.7 Mev. of energy. From the knowledge available at present it is inferred that stars, whose interiors of the sun, obtain their energy by the proton-proton cycle, while stars whose interiors are hotter than the core of the sun produce their energy by another cycle of reactions called the carbon-nitrogen cycle. Estimates of hydrogen in the sun suggest that there is enough hydrogen in the sun to keep the proton-proton cycle going for about 30 billion years.

The energy liberated by nuclear fusion is known as *thermo-nuclear energy*. Very high temperatures and densities are needed for fusion to take place. Because of high temperature, the light nuclei have sufficient energy to overcome the forces of electrostatic repulsion and come together. Because of high density, collisions take place frequently. These conditions for fusion to take place are available in the sun and stars. For the proton-proton cycle to take place large reacting masses such as those we come across in the sun and stars are necessary. There may be considerable time interval between the initial fusion of two protons and their final incorporation in a α particle.

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*Fast moving electrons such as β particles from radio-active materials produce light within a transparent medium if their velocity is greater than the velocity of light in that medium. This in known as Cerenkov radiation.

To adopt the fusion process, for production of energy on the surface of the earth, the reacting mass must be limited in size. Hydrogen bombs were made and tested on the earth. They are fusion bombs. The high temperature, needed to initiate the fusion reaction in a hydrogen bomb, is produced by the explosion of an atom bomb. There are many practical problems to be solved to achieve the production of power by fusion for peaceful purposes.

The fossil fuels (coal and mineral oils), which we use for power generation, will be exhausted after a few centuries. Fission presents a far greater energy source than the fossil fuels. Yet, there are certain disadvantages in the use of fission for power generation. Fission produces radioactive wastes the disposal of which is a serious problem. They can cause great harm to man if they are not properly disposed off. On the other hand, the products of fusion are almost harmless. Secondly, the fuel needed is available in plenty and may last for billions of years. In view of these advantages over fission, research on fusion is going on.

19.6 Radiation Hazards

After the invention of the atomic bomb, studies in the development of nuclear energy have gained utmost importance in many advanced countries. At the present time mankind is faced with the threat of atomic warfare. It is, therefore, important to know the extent of harm that penetrating high energy radiations can cause to human organism.

When a gamma ray or high energy nuclear particle passes through any material it ionizes the atoms it encounters on the way by knocking electrons from them. When an atom, in a complex organic molecule, is ionized the molecule may break up. The breaking up of some molecule could lead to the disruption of the normal functioning of the biological system. This might eventually result in the death of the organism. So, it is the degree of ionization produced, on passing through matter, that determines the strength of the radiation.

The dose of radiation that produces 1.6×10^{12} pairs of ions in 1g of air is accepted as the unit and it is known as the roentgen (r). The number of ions produced in living tissue by a dose of radiation of 1 roentgen is also of the same order of magnitude. One thousandth of the roentgen is the milliroentgen (mr).

The damage to the human organism, caused by penetrating radiation, depends upon (1) the dose, (2) the dose rate, and (3) the part of the body exposed. Hands and feet can receive a much larger dose of radiatoin without permanent injury than any other part of the human body.

The damage can be of two types :

(1) Pathological damage, and

(2) Genetical damage.

Pathological damage to the organism exposed to large doses of radiation could lead to the death of the organism. This will happen if a dose of radiation of about 600 roentgens is delivered over the entire body. Smaller doses of about 100 roentgens may produce leukemia and cancer which will ultimately lead to the death of the organism.

When nuclear radiations and X-rays pass through genetic cells they affect the chromosomes of the cellular nuclei. Injury to the genes in the reproductive cells is serious since it gives rise to mutations which are transmitted from generation to generation. Mutations are almost always harmful and the process is irreversible. It may be noted that, so far as inheritance of genetic damage is concerned, there is no safe lower limit of radiation. Only the probability of absorption of radiation by genes is small when the dose of radiation is small

It is at present accepted that the safe limit for those exposed to radiation over the whole body is 250 mr/week. We are subject to radio-active radiations from the ground and by cosmic rays. The total amount of natural radiation received by a person from the time of conception till the age of forty is estimated to be about 4 roentgens. The additional radiation received during this period from medical X-rays, etc , may be another 4 roentgens. Due to this radiation, the resulting direct damage is negligible because the body has time to repair the injury.

At present the radiation due to radioactive fall-out caused by a atomic bomb tests is found to be only about 0.1 roentgen per person per year. This is small and if no more atomic bomb tests are going to be made. This dose rate will gradually decrease.

19.7 Accelerators

Soon after his discovery of artificial transmutation of elements, using fast moving α particles (See 18 8) Rutherford suggested to Cockroft and Walton to effect artifical transmutation in light nuclei using protons accelerated to very high voltages. A few years later they built a cascade generator for high voltages Using these high voltages they accelerated protons and with them produced artificial transmutation in lithium nucleus.

Van de Graaff Generator

R van de Graaff extended the work of Cockroft and Walton by developing in 1931 the electrostatic accelerator which is called

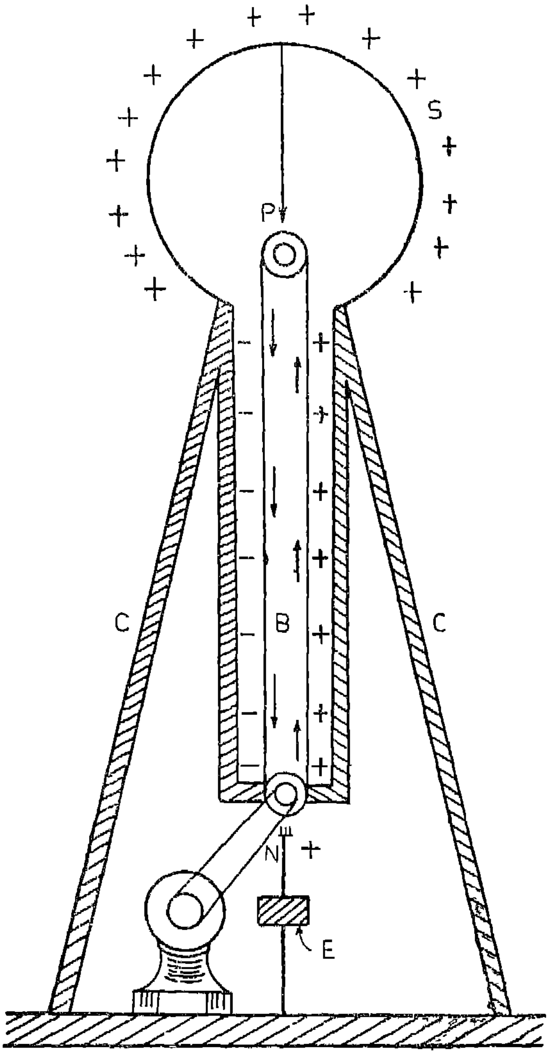

Fig. 19.6 Sketch of a Van de Graff generator,

by his name. This works on the principle that an electric charge imparted to a spherical conductor is distributed entirely on its outer surface

The Van de Graaff generator consists of a large hollow sphere S supported on insulating columns C,C (Fig. 19.6). A conveyor belt B conveys electrical charges from a battery at ground potential to the inside of the sphere. As the belt moves electrons jump from the belt to the row of needle points at N leaving the belt positively charged After entering the sphere the electrons jump from the points P on to the belt leaving the outer surface of the sphere-positively charged. As more and more electrons are removed from the sphere positive potential rises higher and higher until a stage is reached when leakage into the surrounding air and through the insulator becomes significant.

By applying the high electric potential developed in this machine to one end of a long vacuum tube having an ion source, particles such as protons electrons or α particles are accelerated to energies of the order of a few million electron volts.

Cyclotron

Very soon it was realised that this method of artificially accelerating atomic particles to high energies has serious practial limitations. Soon afterwards a new concept of accelerating particles in circular motion, by application of smaller voltages, in repeated steps was conceived and this new type of accelerator known as cyclotron was designed and built by E.O. Lawrence at Berkeley, California University. In this device charged atomic particles are subjected to cyclic or circular motion in a chamber by applying a uniform magnetic field and the particles are accelerated once every half rotation by applying voltage of the order of a few hundred kilovolts across two cylindirical discs known as Dees.

The main part of the instrument consists of two short, hollow, half-cylinders D_1 and D_2 (Fig 19 17). They are referred to as 'Dees' since they are in the shape of D

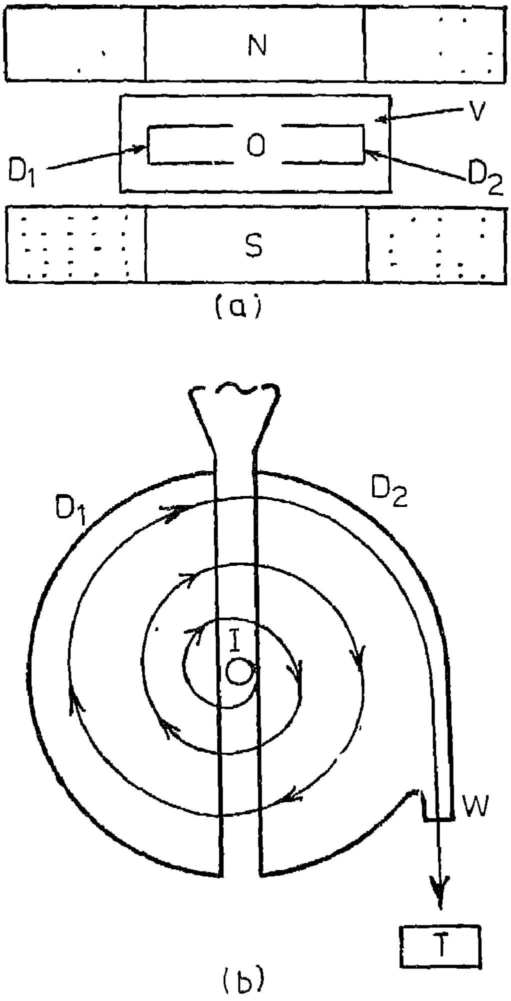

Fig. 19.7 (a) Section diagram of a cyclotron·
(b) 'dees' of a cyclotron showing path of a particle. Main parts of a cyclotron.

They are mounted inside a vacuum chamber V between the poles of a powerful electromagnet. The 'dees' are connected to the two terminals of a high frequency oscillator which supplies the energy to the dees The entire device is in a high vacuum so that the molecules of air will not interfere with the acceleration of the particles.

Let us consider protons, for example, projected into D_1 by means of an ion gun I that works on the principle of the canal-ray tube. Since the particles move at right angles to the magnetic field they take up a circular path. Let the ions after moving through D_1 reach the space between D_1 and D_2. Suppose at this instant D_1 is positive and D_2 is negative. The electric field in the space between D_1 and D_2 accelerates the proton towards D_2 The accelerated proton moves faster through D_2 describing a larger semi-circle than before As the particle just leaves D_2 the polarity of the Dees is automatically reversed by a properly tuned high frequency electronic oscillator. As such D_2 becomes positive and D_1 negative The particle, therefore, receives a second push by the electric field in the space between the dees. It then moves through D_1 describing a larger semi-circle than before as it has now greater energy. Thus every time the proton reaches the gap between the dees the electric field is reversed automatically and the particle finally acquires high enough energy from the many small accelerated pushes it receives. The particle thus moves in a nearly spiral path as shown in (Fig. 19 7 (b)) and will be ejected through the window W by a deflecting voltage and hits the target T.

Theory . The particle moves in a circular path at right angles to the magnetic field Since the magnetic field provides the necessary centripetal force, we have

$$Bev = \frac{mv^2}{R}$$

where B is the magnetic field, e is the charge on the particle, m its mass, v its velocity and R the radius of the circular path

The time t for half a rotation is given by

$$\therefore v = \frac{e}{m} . BR$$

$$t = \frac{\pi R}{v} = \frac{m}{e} . \frac{\pi}{B}$$

t is, therefore, independent of both v and R. This fact is made use of in the above arrangement by making the time period of the high frequency alternating voltage equal to 2t

By changing the frequency of the potential applied to the Dees and adjusting the magnetic field protons, deuterons and α particles of high energies can be produced.

Limitations : The expression for the time t taken by the particle to describe a semi-circle depends upon the value of m, the mass of the particle. We know, from Einstein's special theory of relativity, that mass m increases as the speed of the particle approaches the speed of light. If we try to produce higher speeds in a cyclotron, because of the change in the value of m, the value of t alters and this will throw the particles out of step with the oscillating field. To overcome this drawback many ingenious variations have been made in the construction of accelerators We have as a result, the synchrotron and other machines. These machines employ changing oscillation frequencies or changing magnetic field strengths or both and can accelerate particles upto a few hundred Mev.

In India we have a variable energy cyclotron set up at Calcutta. This is designed to give 6 to 60 Mev Protons, 12 to 65 Mev deuterons and 25 to 130 Mev α particles.

This accelerator will be useful for studies in nuclear physics, solid state physics and radiation biology.

19.8 Cosmic Rays

Around the beginning of the present century, is was noticed that a charged gold leaf electroscope loses its charge even though it is mounted on an insulated support in a closed vessel This was considered to be due to the conductivity of air caused by some unknown agency Experiments proved that this was not due to radio-active materials of the earth. Even on lowering the electroscope into a lake it was found that the rate of discharge of the electroscope remained the same. Experiments using electroscopes carried in balloons upto 25 Km showed that the rate of discharge of the electroscope increased with altitude. In order to explain these observations Hess proposed a new hypothesis that the discharge of the electroscope is caused by a most penetrating radiation falling upon the earth from outside the earth. As the effect is shown both during the day and night, he concluded that this is not part of radiation from the sun but from outer cosmic space These rays were, therefore, named as *Cosmic Rays.*

The gold leaf electroscope and other electrometers were used extensively in earlier studies of cosmic rays upto 1930 They were later replaced by the Wilson cloud chamber and the Geiger-Muller tube. In recent times photographic plates with emulsions specially sensitized for cosmic ray research have been used extensively, particularly for the study of cosmic rays at very high altitudes.

Primary Cosmic Rays and Secondary Cosmic Rays

Experiments have revealed that the cosmic rays entering our atmosphere are almost entirely composed of positively charged atomic nuclei. They are called *Primary Cosmic Rays.* About 89% of these primary cosmic rays are protons, 9% are rays and the rest are heavier nuclei like carbon, nitrogen, oxygen and iron. A high energy primary particle, after entering the atmosphere, soon collides with another atomic nucleus splitting one or both particles into a number of smaller nuclear fragments These fragments share the high energy of the primary particle. They in turn, moving with high speed, collide with other nuclei and produce high-speed particles. In these collisions some elementary particles which are short-lived are also produced During their decay some emit electrons and highly penetrating gamma rays. These photons and all the other high-speed particles that are produced, with the exception of the primary particle, are called *Secondary Cosmic Rays.*

At a height of about 25 km above the surface of the earth the number of secondary cosmic rays is about ten to fifteen times the number of primary cosmic rays that have entered the atmosphere As they move downwards many of the secondaries are stopped by collision because they lose energy by successive collisions This energy is used up in heating the air molecules. A few high speed secondary cosmic rays reach the earth's surface Some of them have enough energy left to penetrate through several hundred metres of earth and water.

Experiments on the latitude variation of cosmic ray intensity showed a minimum near the magnetic equator The cosmic ray intensity at sea level is found to gradually increase on either side of the magnetic equator until a magnetic latitude of 42° is reached and thereafter it remains constant. This is due to the charged particles in the cosmic rays being deflected by the earth's magnetic field. This does not mean that there is no cosmic radiation at the magnetic equator. Only a few high energy cosmic rays enter the magnetic equator.

Cosmic-ray showers . In 1928 Skobeltsyn, engaged in cosmic ray research photographed tracks in a Wilson cloud chamber. The tracks appeared in groups and all the particles in a group seemed to originate from some localized region near the wall of the cloud chamber. These are called *cosmic-ray showers.*

Fig. 19.8 shows the sketch of a cascade shower of cosmic rays photographed in a Wilson cloud chamber. There are some equally spaced lead plates across the cloud chamber A high energy cosmic ray particles entered the solid block of matter and by collission with the nuclei of the atoms in the lead block secondary cosmic rays are produced. The tracks are the paths of these secondaries emerging out of the lower face of the metal block.

While studying the tracks of cosmic rays in a Wilson cloud chamber, Anderson discovered in 1932 a particle having the same mass as an electron but with an equal positive charge which is named as Positron. A positron does not exist for a long period in the free state. When a positron meets with an electron the two are annihilated. Positron is called an *Antiparticle* of electron.

Soon after the discovery of the positron, theoretical physicists began to work out the conditions under which a positron exists in nature. On the basis of the quantum theory of the electron they predicted that, if a high-energy photon (i.e a high frequency gamma ray) were to come close to the nucleus of an atom, the gamma ray would be converted into a pair of particles, namely electron and positron. This is known as *pair production.* The experimental physicists in the field soon found such pairs of particles in a cloud chamber when Gamma rays were passed through matter.

Anderson and Nedermeyer in 1933

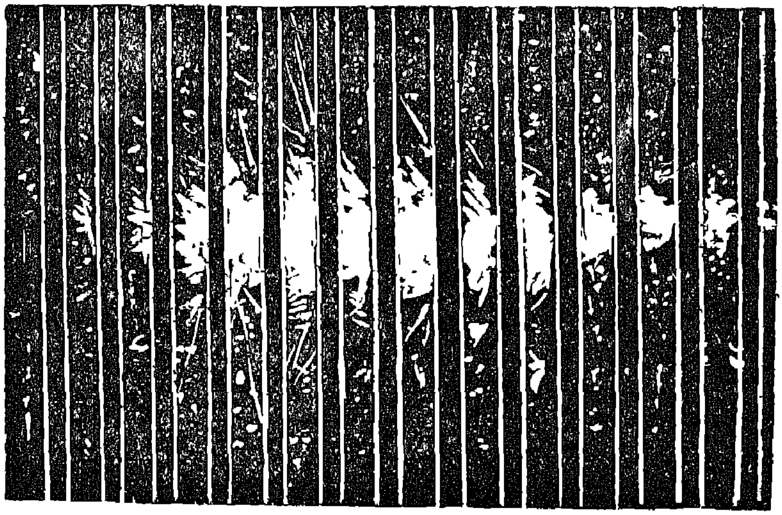

Fig. 19.8 Sketch of cascade shower of cosmic rays,

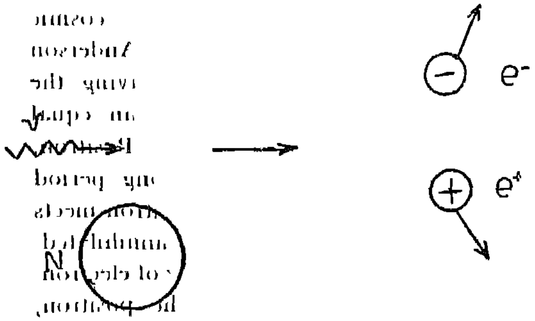

Fig. 19.9 Pair production—a high energy gamma ray comes close to the nucleus of an atom and is converted into an electron and a positron.

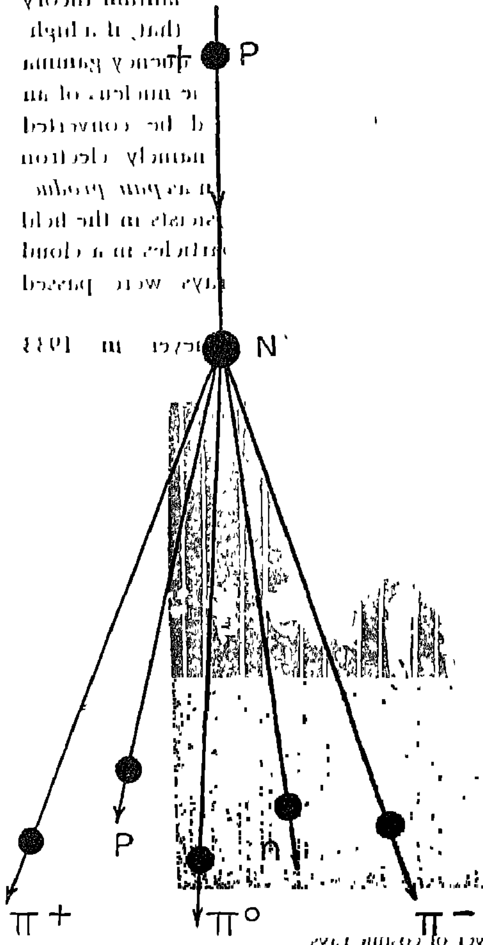

Fig. 19 10 A primary cosmic ray produced, by nuclear Collision, π mesons, protons and neutrons.

discovered the presence, in cosmic rays, of charged particles having mass about two hundred times that of an electron. They are produced by the collisions of cosmic rays with air nuclei. These particles are called mesons (See Section 19.9). They were first predicted theoretically by Yukawa in 1935.

Fig 19.10 shows the production of π mesons or pions, protons and neutrons by a primary cosmic ray by nuclear collision. Some of the pions are positively charged, some negatively charged while some others are neutral.

All charged pions spontaneously disintegrate into a μ *meson* or muon and a *neutrino* (Fig. 19 11). The negative muon decays into an electron and two neutrinos. Neutral pions (π° mesons) are without charge and very unstable. They decay into two gamma rays. By electron pair-production these gamma rays create cascade showers of electrons in the upper regions of the atmosphere (Fig. 19.11).

19.9 Elementary Particles

From early times, natural philosphers believed that all material things are made up of a few basic particles. In their search for these fundamental or elementary particles, physicists have used atomic projectiles from powerful accelerators to bombard atomic nuclei. At the time of the discovery of the neutron, only four elementary particles were known. They are the electron, proton, photon and neutron. Of these particles, the electron, proton and neutron have definite rest mass but the photon has no rest mass. Subsequently more and more elementary particles were discovered. Some of them were predicted theoretically and confirmed later by experiment.

In radio-active decay it was found that the laws of conservation of momentum and

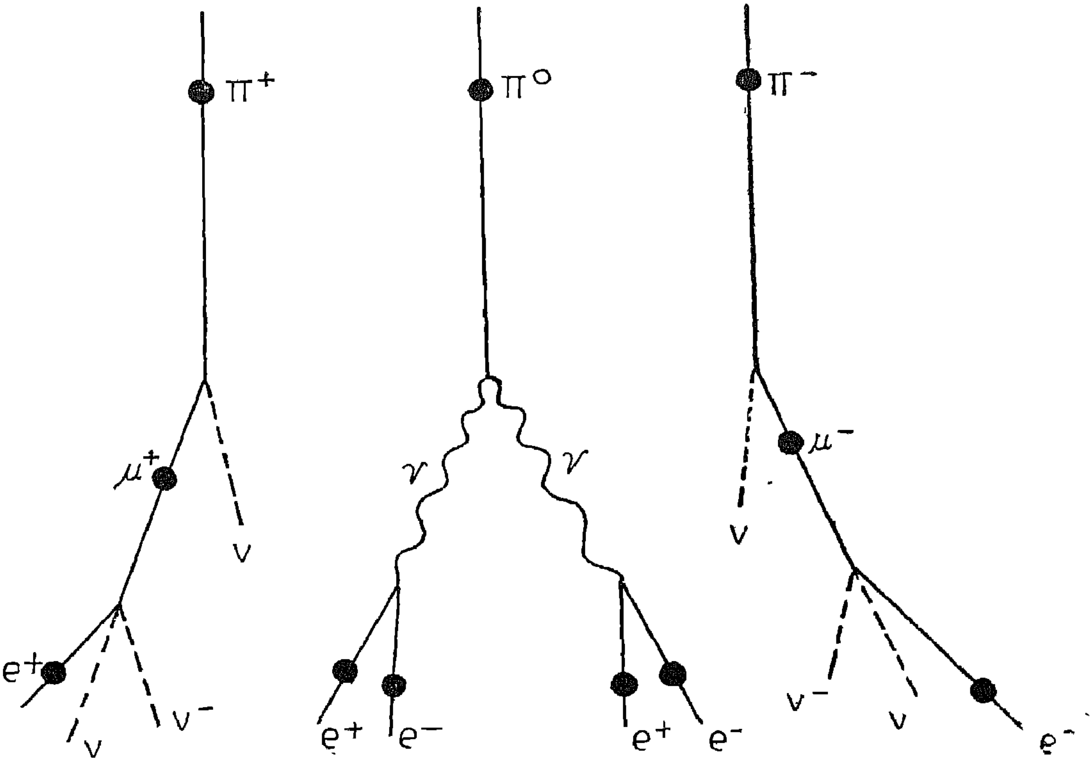

Fig. 19.11 Disintegration of π mesons into μ mesons, α rays, electrons and neutrons.

energy were obeyed in the case of alpha and gamma decay. On the other hand, in beta decay a serious discrepancy was observed in that the laws of conservation of momentum and energy were not found to be obeyed. This discrepancy led pauli to postulate the existence of a new particle named neutrino which has no mass and charge but has angular momentum and energy. At the same time, it does not have the characteristics of an electromagnetic wave like photon. By postulating the emission of a neutrino along with the electron in beta decay process (Fig.19.12), it was observed that the laws of conservation of momentum and energy would be obeyed. The existence of the neutrino was experimentally proved in the year 1956. Since neutrino is a particle without any charge or mass it is capable of penetrating through large thickness of matter without appreciable interaction. They are generated in great abundance in the sun and in nuclear reactors. In beta decay in which a positron is emitted a neutrino is released, where as in beta decay involving electron emission an antineutrino is emitted, as indicated below :

$$p \rightarrow n + e^{+} + \qquad + \upsilon$$

$$n \rightarrow p + e' + \qquad \upsilon^{-}$$

Earlier, in cosmic ray studies, reference

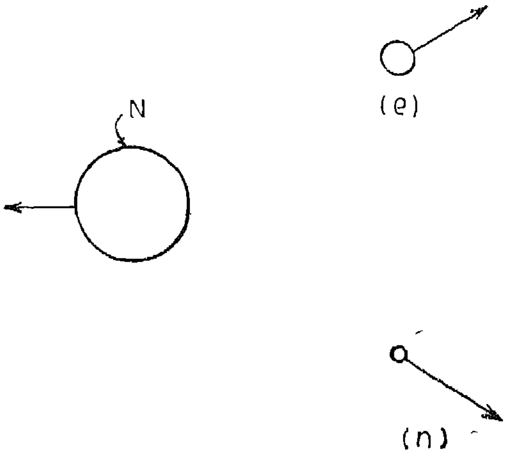

Fig. 19.12 Emission of electron (e) and neutrino (n) during beta decay, N—nucleus after β-decay.

was made to the discovery of the position which is an antiparticle of the electron. When a positron comes into contact with an electron both get annihilated resulting in the emission of two gamma rays (Fig. 19.13) each having an energy of 0.51 Mev.

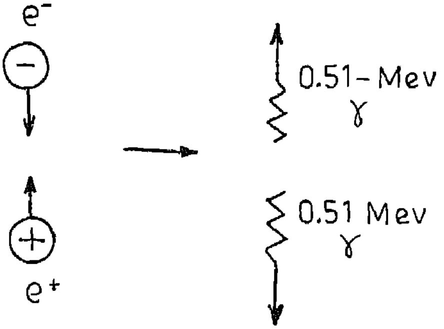

Fig. 19.13 Annihilation of positron and electron, when they come together, producing two gamma rays each having 0 51 Mev enery.

After the discovery of the positron the existence of a negative counterpart of proton was also predicted. Such a particle called the antiproton was experimentally discovered much later. When an antiproton comes into contact with a proton both are annihilated resulting in the production of a π meson with large energy Sometimes instead of a π meson, two gamma ray photons, each having 936 Mev energy, are produced (Fig 19.14).

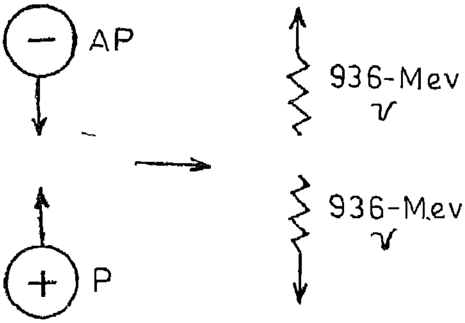

Fig. 19.14 Annihilation of a proton and anti-proton when they come together producing two gamma rays each having 936 Mev energy.

We have referred, in the section on cosmic rays, that two types of mesons, namely, π mesons or pions and mesons or muons were discovered in cosmic ray studies. A charged π meson has a mass of 273 m_e. Its life-time is 2.54 × 10^{-8} seconds. A charged π meson decays into a meson and a neutrino as indicated below :

$$\pi^+ \rightarrow \mu^+ + \nu$$

$$\pi^- \rightarrow \mu^- + \nu^-$$

There is also a neutral π meson which decays into two gamma rays

$$\pi^0 \rightarrow \gamma + \gamma$$

The μ meson has a mass of 207 m_e and has a life-time of 2.22 × 10^{-6} seconds. The muons have great penetrating power because of their feeble interaction with matter. A muon decays giving an electron and two neutrinos as indicated below:

$$\mu^+ \rightarrow e^+ + \upsilon + \bar{\upsilon}$$

$$\mu^- \rightarrow e^- + \upsilon + \bar{\upsilon}$$

There is no neutral μ meson. Mesons are also classified as elementary particles.

Although, at one time, the total number of elementary particles was about thirty, recent researches have added more elementary particles to this list increasing the number to above 200. As the research work on elementary particles is advancing more and more, new particles with increasing complexity are being discovered.

Exercises

19.1 The isotope uranium ${}_{92}U^{238}$ decays successively to form ${}_{90}Th^{234}$, ${}_{91}Pa^{234}$ ${}_{92}U^{234}$, ${}_{90}Th^{230}$, and ${}_{88}Ra^{226}$. What are the radiations emitted in these five steps ($\alpha,\beta,\beta,\alpha,\alpha$)

19.2 If 200 Mev energy is released in the fission of a single nucleus of ${}_{92}U^{235}$, how many fissions must occur per second to produce a power of 1kw? (3.12×10^{13})

19.3 Give an outline of the working of a simple type of nuclear reactor.

19.4 Assuming that in a star three alpha particles join in a single fusion reaction to form a ${}_{6}C^{12}$ nucleus calculate the energy released in this reaction, given mass of ${}_{2}He^{4}$ = 4 002 604 amu and that of ${}_{6}C^{12}$ = 12.000 000 amu. (7.3 Mev)

19.5 Assuming that four hydrogen atoms combine to form a helium atom and two positrons, each of mass 0 000 549 amu calculate the energy released. (25.7 Mev)

19.6 Write a note on radiation hazards.

19.7 Describe the construction and explain the working of a van de Graaff generator.

19.8 Deuteron are to be accelerated in a cyclotron. The value of the uniform magnetic field applied is 2.475 Wb/m². What is frequency of the oscillating potential that must be applied to the dees of the cyclotron? (mass of deuteron = 3.3×10^{-27} kg and charge 1.6×10^{-19}C) (19.0 H_2)

19.9 What is the period of revolution of a deuteron in a uniform magnetic field of 1 Wb/m². Use the data given in question no. 8. (1.296×10^{-75})

19 10 What are the primary Cosmic rays composed of ? What are the secondary Cosmic rays composed of ?

19.11 Why is the intensity of cosmic rays a minimum near the magnetic equator?

19.12 Draw diagrams to show the disintegration of Pions.

19.13 What is the antiparticle of the antineutron?

19.14 How much energy must a gamma ray photon have if it is to materialize into a proton-antiproton pair with each particle having a kinetic energy of 10 Mev? (mass of ${}_1p^1$ = 1.007 276 amu) (1896 Mev)

19.15 An electron positron pair is produced by a gamma ray of 2.26 Mev. How much kinetic energy is imparted to each of the charged particles ? (0 62 Mev)

19.16 A 1 Mev positron collides head on with a 1 Mev electron and they are annihilated giving rise to gamma ray photons. What is the total energy of the resulting gamma rays?
(mass of electron 0.000 549 amu) (3.02 Mev)

UNIT 20

Electronic Devices

We live in an age of technology and one branch of technology that has revolutionised the present civilization is, of course, electronics. It is now possible for people to communicate speedily and efficiently with each other by using electronic appliances, such as radio, television, etc which are the products of this technology Similarly, radar and sonar and many other electronic instruments are used to defend our country from others. There are many other examples in our every day life that can be cited to show that electronics has now become an important tool of progress and welfare of the modern society.

Most interesting feature of all electronic devices is their speed of action. For example, the conventional method of communication between two individuals by letters etc takes longer time in comparison with the almost instantaneous communication that can be achieved with the help of radio waves. Similarly, the time taken to multiply six figured numbers by conventional procedures may take many seconds while it takes only a few microseconds to do the same multiplication by an electronic calculator. It is because of such speeds, now achievable with electronics, that this technology has become so important.

20.1 Essentials of an Electronic Device

Electronics is essentially a branch of science which deals with use of electrons for specific applications Electron as we know is the fundamental particle with the smallest mass and finite electrical charge. It is due to these two properties that electrons can be speedily and easily manipulated. Manipulation of electrons can be done either in the form of their number or their positions. We have already studied the effect of magnetic and electric fields on electrons and are quite aware that by applying such fields it is possible to change the direction of their motion conveniently. Moreover, the energy that is required to manipulate this way is very small. At the same time, the velocities attained by the electron in the process are very large and hence account for high speed of action. All electronic devices essentially make use of these important properties of electrons.

It is now clear from the above discussion that electronic devices must therefore have the following three parts :

1. a source of electrons
2. an arrangement for manipulating them, and
3. an arrangement for observing the

desired effect due to the manipulation.

In this unit we will see that these three features are essential to all electronic devices.

20.2 Source of Electrons

Since electrons form a major constituent of atoms which constitute the matter they are available in large numbers. They however, tend to interact with matter that comes in their contact due to their peculiar electrical nature ; and form bonds. This tendency, in fact, accounts for the difficulties in freeing them from substances, particularly, the insulating materials which are the most abundant form of matter around us. It is, therefore, necessary to look around for these materials in which the binding of electrons is weak enough to allow for easy removal for further use from their bonding atoms.

(*a*) *Metals as source of electrons*

In the chapter on solids we can see that metals contain freely moving electrons which account for the high electrical conductivities of the metals. These electrons are, however, confined to the conductor and cannot leave its surface at the nominal temperatures or electric fields. It is easy to understand if we consider what happens at the moment, the electron decides to leave the surface. The removal of the electrons immediately produces positive charges on the surface of the conductor which was electrically neutral before the electron came, just out of the surface. The positive charge therefore, pulls back the electrons that try to esacpe. Thus the electrons in the metal which happen to come on the surface do not get opportunity to leave and hence, remain virtually confined to it. This situation is very similar to that of molecules of water trying to escape against forces at the surface in the process of evaporation. In the case of electrons, however, the pull at the surface is governed by the nature of the material. The energy that is required to remove the electrons from the surface so as to overcome the retaining forces at the surface, is called the work function and is measured in electron volts (eV). Typical values of work function for various materials are given in Table 9.2. It is seen that the energy needed to remove electrons from tungsten and molybdenum is higher than that required for their removal from barium and cesium.

It is now obvious that if we choose to use electrons which are easily available in metals, some methods of overcoming the pull at the surface must be devised. One method which can be easily thought of is, of course, to subject the metal to a very high electric potential which will extract the electrons from within and make them free This type of removal of electrons by applying high electric fields is called field emission. Since the use of high voltages is technically difficult and often dangerous, the field emission is not used as a method of obtaining electrons in appliances often used in the home.

While studying the physics of atoms, we have already studied photoelectric effect and the laws of photoelectric emission Photoemission from materials can also act as a source of electrons and has been conveniently used in some special devices. In this case, however, the energy of photons required to release the electrons from the metal surface must be greater than the work function. Both the photo emission and the field emission have limited applications because the rate of removal of electrons, i.e., the current available from these sources is generally very small.

There is another method of getting electrons in large numbers and this has been in use for a number of years. In this method a metal filament is heated by passing enough current through it. Increase in the temperature of the filament increases the thermal energy of the electrons which in turn helps them to overcome the pull at the surface. At sufficiently high temperatures, a reasonably high number of electrons find themselves with energies larger than the work function and leave the filament. This type of emission is called the *thermionic emission*

The electrons emitted from the surface of the filament are not directly available for manipulation but tend to remain more or less in the space near the filament in the absence of an electric field, and form a cloud known as *space charge.* Since like charges repel each other, this cloud of electrons tends to force back to the filament other electrons that are being emitted The amount of electron emission is thus governed by the space charge that the cathode has around it.

It is obvious that, in order to obtain substantial emission of electrons, the metal should be heated to as high a temperature as possible. Therefore, it is customary to use metals having high melting points such as tungsten or molybdenum as materials for heaters. Being metals themselves, these heaters can provide emission of electrons However, their work functions being higher, they are not efficient enough for large emission currents. In order to improve the efficiency of emission, tungsten heaters are generally coated by materials like barium oxide which have low work functions. Such high efficiency heaters are used in vacuum tubes

(*b*) *Semiconductors as source of electrons*

So far we have seen that to obtain electrons for devices, we need to remove them from the metals by various techniques, into the free space for manipulation. These methods have technical disadvantages such as operation at high temperature, provision of energy for heating etc Moreover, since the electrons are free, they have to be manipulated for the purpose of the device only in an evacuated space, where they would not collide with gas particles which would only make manipulation difficult. The device therefore becomes bulky.

Since the last three decades we have been using devices made of semiconductors which do not have all these unfavourable characteristics These devices are small and do not require additional heating, and hence can work at room temperatures This is possible because semiconducting materials can provide free electrons which can be manipulated in the material itself without bringing them into the free space. Thus these devices are very small, hardly a square mm in size. We have already studied in the Unit on Solids, the way electrons behave in semiconductors. Here we will consider only those aspects which are relevant to electronic devices.

Germanium and silicon are the semiconducting elements most commonly used in devices. In the crystalline form, their atoms are placed in the diamond like structure, and bonded covalently with each other. The bonding electrons i.e., the outermost electrons of these atoms, are bound with so small an energy that it is possible even at normal room temperatures to free some of them from their atoms. Once freed they can move about easily until caught by another atom which had already lost its electron The number freed, however, will depend on their bond energy in the semiconductor. For example, silicon which required higher energy for dislodging an electron from its bond gives less number of free electrons for a given temperature compared to germanium which has correspondingly lower bond energy. This accounts for the fact that majority of semiconducting devices are made of silicon which is least affected by temperature.

Let us now see what happens if we apply electric voltage across such a semiconducting crystal. Fig 20 1 shows a piece of germanium connected to a battery, a switch and an ammeter to record current that would flow in the circuit In fact, when the switch is put 'on' the meter would show current flowing through the circuit. This current most obviously be due to charges moving in the semiconductor. This is hardly surprising, considering the fact, that the experiment is done at room temperature i.e., about 300K which provides enough thermal energy to release electrons and move under the influence of the applied electric field and register a current. We have already learnt in the unit that this current is composed of two opposite currents produced by 'n' type carriers, i.e., electrons and 'p' type carriers i e, holes. They are produced in equal number to account for the electrical neutrality of the material. This type of material, which has to be very pure, is called intrinsic semiconductor. If we increase the temperature of the semiconducting piece, say, by holding it over a flame, current in the meter is seen increasing very rapidly When the semiconductor is cooled to the room temperature, the current drops to its original value. Those charge carriers can also be produced by exposing the semiconductors to light, however, unlike photoemission in metals, in which it escapes the metal surface, the electron released by the striking light remains within the semiconductor and hence can be easily manipulated. For an efficient device, it is necessary to have a source of electrons which can give ample but controllable number of electrons. Intrinsic semiconductors, in that respect prove to be a poor source of electrons, and technically difficult to control or manipulate. Instead, by adding impurities of the right type, called dopants, at the time of preparing crystals of semiconductors it is possible to enhance the number of either 'n' or 'p' type charge carriers in them. Semiconductor that gives 'n' type carriers is called an 'N' type semiconductor while that which gives 'p' type charge carriers is called a 'P' type semiconductor. It is easy to see that by controlling the amount of dopant, we can control the strength of our sources of charge carriers and fabricate them to suit the requirement of the device to be made.

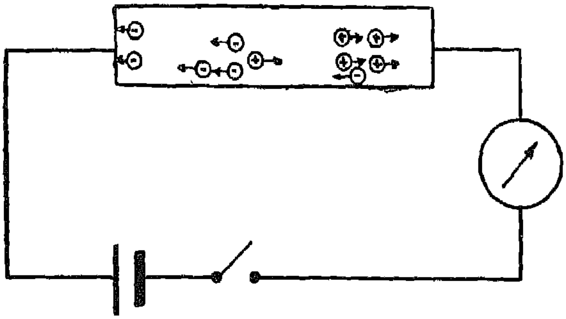

Fig. 20.1 Semiconducting Germanium crystal connected to a battery

It is not always necessary in practice, to use only doped crystals. Since the size of the actual crystal is small, the dopants can be made to enter the desired regions of a pure crystal by depositing them on the surface and diffuse the dopant atoms at higher temperatures. This makes fabrication of semiconducting devices rather simple.

20.3 Manipulation of Electrons

So far we have studied various sources of electrons or charge carriers which can be used for making electronic devices. As has been mentioned earlier, these electrons are manipulated in the device to produce the desired effect. It is, therefore, obvious that the choice of method of manipulating electrons will depend mainly upon the final effect we are looking for. In this section therefore we will try to recapitulate some of the methods of manipulating electrons which have been studied earlier in the course.

There are two ways of manipulating electrons One simple way is obviously to change their numbers, while in the other the number of electrons is kept fixed but their direction of motion in the device is changed to produce the desired effect. The first method involving the number of electrons is easy to understand because we know that in most of the electrical devices, currents can be increased by increasing the voltage applied to the device and vice-versa. For example, electrons when flow through a resistance will give rise to a current as a result of the voltage applied to it and according to the Ohm's law if the voltage is doubled the current is also doubled and so on. Therefore, control on the number of electrons can be achieved by merely subjecting the device to appropriate voltages. Such devices are called linear devices because they behave according to a linear relation such as Ohm's law.

The second method of manipulation, however, depends on changing the original path of the electron to make it reach a different position so that the desired effect can be obtained. We have already learnt that if a beam of moving electrons is subjected to an electric field, acting perpendicular to the path of the electrons, then it deflects in the direction of the electric field The resultant displacement is decided by the force acting on the electrons in the direction of the field and its original velocity Thus by changing the field in a desired manner, it is possible to manipulate the position of an electron The other method of deflecting electron beam in space is by subjecting it to a magnetic field again perpendicular to the original direction of motion. The resultant shift in the path will then be governed by Lenz's law. These two methods of manipulating the positions of electrons are effectively used in devices such as television tubes and oscilloscope tubes.

20.4 Diode

Now that we know the basic elements of any electronic device, namely the source of electrons and the methods of manipulation, we are in a better position to understand their functions.

The simplest type of electronic device constitutes an electron emitting source on oxide coated tungsten filament, called a cathode, and a surrounding metal electrode to collect electrons called anode or plate, both enclosed in an evacuated envelope. Such a device is called a *diode* (*di* means two in Greek and *ode* is from the word electrode). The structure of a diode is shown in Fig. 20.2. We have already learnt that the diodes are used as elements to rectify a c. voltage i.e., to convert a.c. voltages into d c. voltages. The process of rectification demands that the diode allows

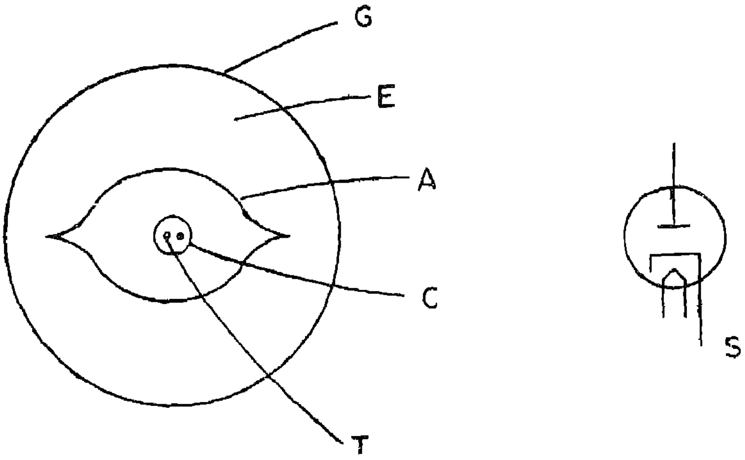

Fig. 20.2 Diode G—glass envelope, E—evacuated space, A—anode, C—cathode, T—tungsten filament, S—symbol for diode.

current to be drawn only in one direction on application of an alternating voltage. To be precise, we can say that when the anode is positive relative to the cathode, the electrons from the cathode should flow in one way only from the cathode to the positive anode On reversing the polarity of the voltage, there would be no flow of electrons. In other words the diode can be considered as a sort of switch which is 'on' when the anode is positive relative to the cathode and opens when the respective polarities are reversed Since electrons are used for such a switching it is an extremely fast acting switch.

(*a*) *Working of a diode*

If a positive voltage is applied to the anode, the electrons from the space charge cloud which are close to it are first attracted giving rise to the current. However, the removal of the electrons from the space charge will allow equal number of electrons to come out of the cathode surface and maintain the number of electrons in the space charge the same. The electrons which were attracted to the charged plate would neutralise the positive charge as shown in the circuit in Fig. 20.3. A current is registered due to the passage of the charge back to the cathode. Internal flow of electrons from the cathode to the anode is called *space current*. In this type of device the anode current is equal to the space current.

The space current as it is easy to see, will flow only in one direction from the anode to the cathode. Since the plate is not a source or emitter of electrons, but merely collects them, there is no charge available to flow from the anode to the cathode in the event of the reversal of the polarity of the voltage.

(*b*) *Diode characteristic*

Let us now see how the anode current behaves when the anode voltage is changed. As the positive voltage on the anode is gradually increased more and more electrons from the space charge are taken up by the anode and therefore the current increases with the voltage. After a certain voltage the space charge is completely removed and the electrons, emitted by the cathode, travel directly towards the anode so that the rate of release of electrons by the cathode is equal to that at which the plate collects them. Increasing anode voltage beyond this point will, however, not result in an increase in the current, since entire emission of the cathode has been taken up by the anode. This limit of current that can be reached is called saturation current. The dependence of anode current on its voltage is shown in Fig. 20 4.

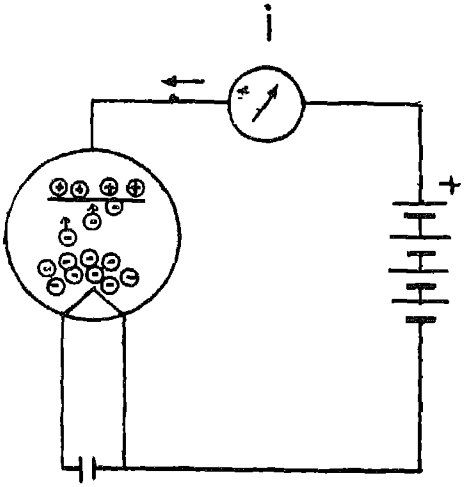

Fig. 20.3 Diode circuit, i—space current.

(*c*) *Diode as a rectifier*

Suppose we now replace a battery in Fig. 20 3 by a generator of a c. voltage The common a c. voltage available from the domestic supply consists of a voltage varying sinusoidally at the rate of 50 times a second. Fig. 20 5(a) shows such a voltage applied to the anode. The a.c petential here swings in one direction for half the cycle and in opposite direction for the remaining half. If we assume that on application of such a voltage, the first half of the cycle alone makes the anode go positive with respect to the cathode, then the current in diode will flow from the anode to the cathode. However, on reversal of polarity due to the later half of the cycle, the diode will not pass any current. The nature of the current is shown in Fig. 20 5.(b), which essentially is a pulsating current, but remains on one side of the zero axis, all the time. It, therefore, can be considered as a direct current (d.c.) with added variation in its amplitude. This process of changing alternating current to direct current is called rectification and in majority of cases, the diodes are used only for this purpose. The rectifying action of the diode in relation to its characteristics can be visualised by using a graph shown in Fig. 20 6(a). Here we have really combined two graphs in one since they are related to

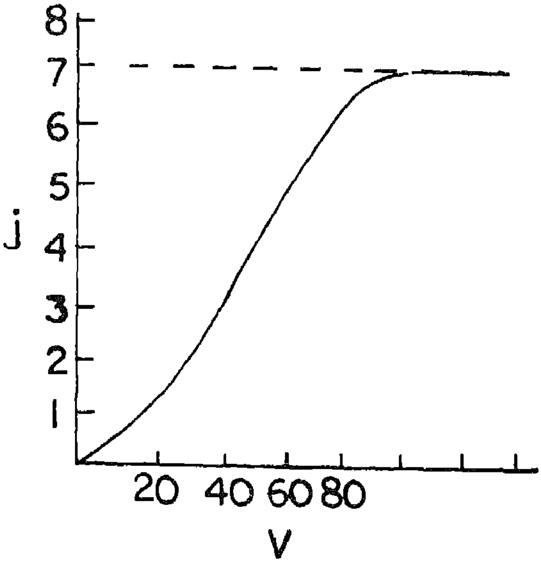

Fig. 20.4 Diode characteristic
V—anode voltage, j—current

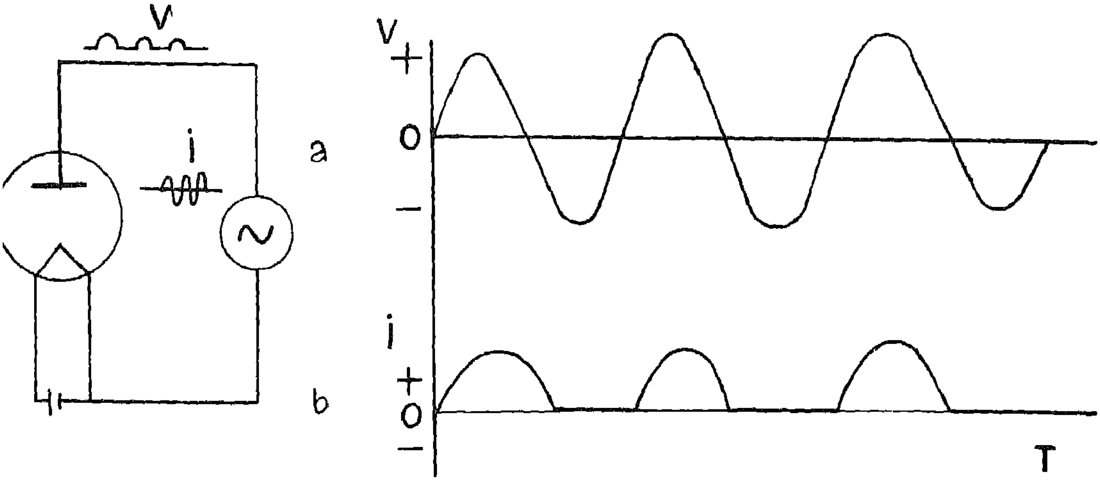

Fig. 20.5 Rectification of A C. Voltage applied to a diode. V—anode voltage, i—anode current, T—time. (b) Rectified current.

each other. The vertical axis here is extended downward to represent time and is used to show sinusoidal variation of the anode voltage. The sinusoidal variation is shown here to have been broken into four equal intervals. On the horizontal axis, we again have time to show the resulting current governed by the diode characteristics indicated by the pulsating d.c. output current which appears only in the positive half cycles There is an advantage in showing the behaviour of the diode in this fashion. It will be obvious that if we apply higher and higher a.c. voltage to the diode at some voltage the diode will reach a point where it cannot deliver more current because of the limits imposed by the saturation current, as shown in Fig. 20.6(b). It is, therefore, obvious that for proper operation of a rectifier it is necessary to stay within the

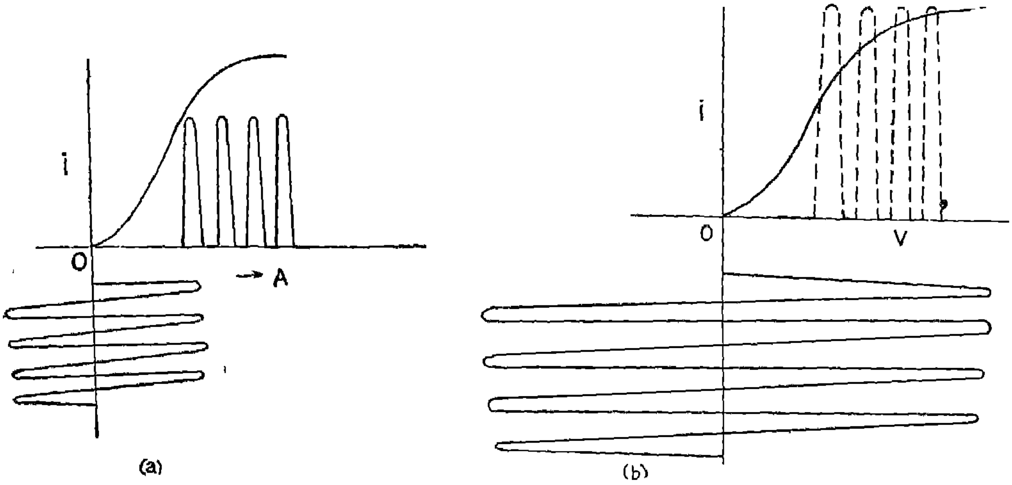

Fig. 20.6 (a) Rectification action of a diode in relation to its characteristics. V—anode voltage i—anode current. (b) Saturation.

limits of voltage specified by the design of the diode. It is also clear that the voltage current relationship of a diode is no more linear and hence such a device will behave as a non-linear device in contrast with a simple resistance which follows the linear Ohm's law

In practice such a pulsating d.c. is of very little use. It is, therefore, necessary that before putting it to use some arrangement must be made to remove the pulsations and extract the d.c. part of this voltage. This is done by smoothing circuits consisting of capacitors and inductors. Fig. 20.7 shows a circuit in which the capacitor acts as a reservoir of charge, delivered to it by the diode, and helps to nearly remove the pulsations in the voltage across the resistance connected across it. This is a simple method of developing d.c. voltages which can be used as a source of power for electronic circuits.

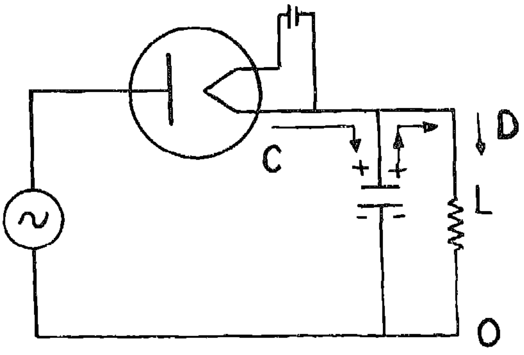

Fig. 20.7 Use of a capacitor in diode circuit C—charging, D—discharging, L—inductor

(d) Diode as a detector

We have already learnt that for long range communication with radio waves it is necessary to modify a sinusoidal radio frequency in such a way that it can carry audio signal on it. This process of impressing a signal upon a carrier is called modula-

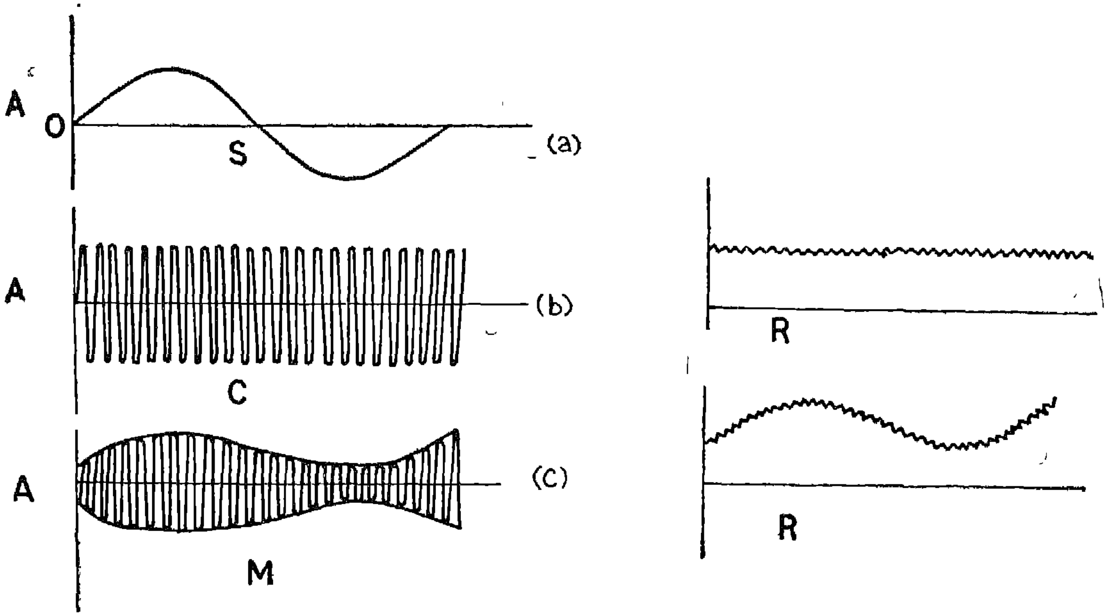

Fig. 20.8 Modulation of a sound signal, S—audio signal, C—carrier wave, M—modulated wave, R—rectification, A—amplitude.

tion Let us consider a relatively simple case of a transmitter, transmitting a signal which consists of a single frequency Fig. 20.8(a). The transmitter will generate a suitable carrier wave as shown in Fig. 20.8(b). In the transmitter the sound signal and the carrier are mixed to produce a modulated radio frequency carrier, by a method in which amplitude of the carrier is made to vary in direct relation to the audio frequency. Such a wave is called amplitude modulated wave and is shown in Fig. 20.8(c). The original audio frequency wave is now on both the sides of the carrier.

Let us now apply the carrier voltage in Fig. 20.8 (b) to a diode rectifier circuit as had been done earlier in Fig. 20.7. Since the amplitudes of all the waves of the carrier are equal, on rectification and proper smoothing, we would obtain a d c voltage proportional to the amplitude of the carrier wave. Suppose that we now apply the voltage of Fig. 20.8 (c) to the same circuit, since the amplitude of each carrier wave cycle is varying in accordance with the impressed audio signal, we will get on rectification, a voltage which will be similar to the audio signal. This is shown in Fig. 20.8. Thus the diode can be used to demodulate the modulated wave or detect the signal from the carrier

20.5 Semiconductor Diodes

So far we considered the devices which used electrons moving from one electrode to another on application of electric field, by making one electrode a source of electric charge and the other a mere collector. We have seen that the current in these devices can flow only in one direction. In the past two decades semiconductor diodes are largely used in electronic circuits. In a normal piece of semiconductor it is not possible to make one part of it behave as a source of charges and another as a collector One has to artificially generate conditions such that one part of the piece acts as a source of electrons while the other would be deficient in electrons and therefore would collect them on application of an electric field.

(*a*) *p-n Junction*

In is only under this condition that such a system will behave in a way similar to a diode. In practice this is achieved by forming a junction between an 'N' type crystal to a 'P' type crystal as shown in Fig. 20.9. It is uneconomical to use two separate crystals to form a junction, instead by a suitable technique, regions having the necessary N or P type of material are created side by side in the same piece of semiconductor, thus forming a p-n junction.

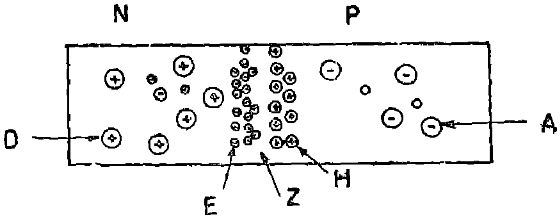

Fig. 20.9 p—n junction N—N region , P—P region, D—ionised donor, A—ionised acceptor, H—hole, Z—depletion zone, E—electron.

Let us now see the behaviour of charge carriers in a p-n junction. In the 'N' region of the crystal we have greater concentration of electrons than that in the 'P' region. Hence some of the excess electrons of the impurity atoms, called the donor atoms, in the 'N' region will drift in to the 'P' region after breaking away from their parent atoms. After losing the electrons the donor atoms in the 'N' region, will become positive ions Since they are fixed in place within the crystal framework at the lattice points, the positive ions thus formed remain fixed in the 'N' region near the junction. The 'N' region near the junction has, therefore, lost some of its electrons and is no longer neutral, it is positively charged. Similarly holes from impurity atoms, called acceptor atoms, in the 'P' region would drift towards the junction giving rise to the negative ions. As a result the 'P' region close to the junction similarly loses electrical neutrality and is negatively charged.

The electrons which have diffused into the junction from the 'N' region meet holes from the 'P' region at the junction. Being oppositely charged, they recombine thus removing both the carriers. However, as more holes from the 'P' region are forced to diffuse across the junction they are repelled by positive ions in the 'N' region Similarly the electrons from the 'N' region which attempt to diffuse across the junction towards the 'P' region are repelled by negative ions of the 'P' region. This fact gives rise to accumulation of -ve charges at the junction on the side of the 'N' region and similar accumulation of +ve charges on the side of the 'P' region with a very thin separation between them. Such an accumulation of opposite charges on two sides of the junction results in creation of an electric field across it, called a barrier potential. It is called so because the charges must overcome a barrier like potential to cross the junction. Its strength is decided by the nature of the original material from which the junction is prepared. For example, in the case of silicon it is of the order of 0.6 volts and somewhat less in the case of germanium. This barrier stops further diffusion of the electrons from the 'N' region and of the holes from the 'P' region. The zone within which this barrier potential is developed is called the 'depletion' zone.

(*b*) *Junction rectifier*

Let us now consider what happens if a p-n junction is subjected to an external electrical field supplied by a battery, as shown in Fig 20 10. The negative terminal of the battery is connected to the 'N' region and the positive terminal to the 'P' region. This arrangement of applying voltage to a junction is called forward bias.

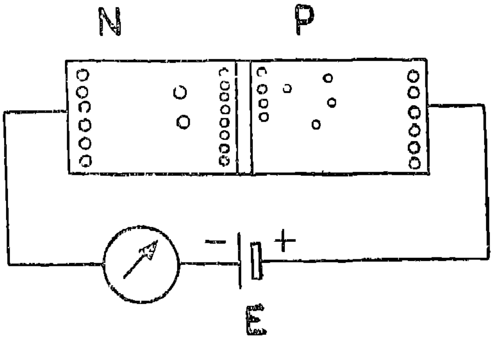

Fig. 20.10 p-n junction connected to a battery (forward bias) N—N region, P—P region.

On application of this electric field, the electrons of the 'N' region are repelled from the negative terminal of the battery and drift towards the junction. Similarly, the holes from the 'P' region would drift towards the junction because of the positive charge at the other end. This will force carriers to move towards each other thus reducing the width of the depletion zone. In addition the external field being in opposition to the barrier potential, the strength of the barrier potential is reduced.

Once the strength of the barrier potential has been reduced the electrons are now able to cross the junction meeting the holes which have crossed the barrier to enter the depletion zone in a similar way and recombine with them. These recombinations of electrons and holes is registered as current in the external circuit. As long as the external battery is applied this flow of charges will continue to take place thus giving rise to the forward current. Although the height of the potential barrier originally was a few tenth of a volt, it is often seen that voltages as high as one or one and half volts are required to overcome the barrier. The voltage excess over the original potential barrier is due to the drop that takes place in the body of the crystal, forming the depletion region. Having once overcome the barriers, charge carriers will flow giving rise to the forward current which will increase with increasing voltage. However, it will be accompanied with an increasing energy loss in the body of the crystal due to the voltage dropped in the other parts of the crystal. Such an energy loss appears as increase in the temperature of the device which is detrimental to the functioning of the diodes. The relationship between the forward currents and the applied voltage for a p-n junction diode is shown in Fig. 20.11. Note here that the current does not start flowing for voltage smaller than that required for overcoming the potential barrier.

Let us connect the external battery in the opposite manner such that the positive terminal is now connected to the 'N' region and the negative terminal to the 'P' region. In this case it is said to have a reverse voltage or bias applied to the p-n junction. A careful observation of the way the charge carriers would now behave shows that the electrons from 'N' region will be attracted towards the positive end of the battery, and the holes will similarly be attracted, towards the negative end of the battery, thus keeping the carriers away from the depletion area. In this case obivously no current will flow, since on application of electric field in this manner, the barrier potential will be more than what it originally was. Therefore, in principle no current should flow in a reverse bias p-n junction. However, at room temperature, there is always a possibility that some of the donors and acceptors which are near the depletion zone give away charges giving rise to a recombination and hence a very small reverse current. In fact we have seen that p-n junction can act in a way similar to the diode.

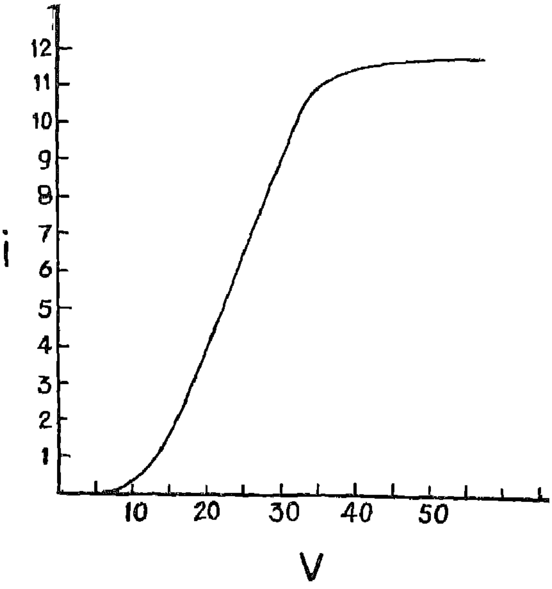

Fig. 20.11 Relationship between forward current and applied voltage for a p-n junction

Such junctions are now extensively used in transistor radios, televisions and many other applications of electronics.

20.6 Triodes

In the diodes so far studied by us, the

manipulation of electrons to produce affects like that of rectification was done merely by controlling the applied external electric fields to its two electrodes. It is possible to have a better control on the electrons by providing an additional electrode for manipulating their numbers. There are two types of devices using this technique of manipulation The first type is known as a triode (*tri* for three and *ode* from electrodes) in which electrons are controlled in vacuum and second is the transistor, which similarly controls charge carriers in semiconductors.

(*a*) *Construction of a vacuum triode*

First we will look into the working of a vacuum triode and see how the manipulation has been done in a more refined fashion for totally different purpose compared to that in the diode Let us insert another electrode in between the cathode and the anode of the diode, after taking care to see that the electrons can pass through the new electrode. It is possible to do so by providing a wire mesh which will permit us to establish the necessary electric field for manipulation but will allow the manipulated electrons to pass through its structure. This electrode, is, therefore, called grid. In actual triode this grid consists of spiral or an open mesh of circular or elliptic cross section and is made of a thin iron, molybdenum or special alloy wire. For efficient control and collection of electrons, the filament or cathode is held at the centre and the grid is kept close to the cathode surrounding it. Both of them are then surrounded by a cylindrical anode for efficient collection of electrons. The structure of the triode is shown in Fig. 20.12.

(*b*) *Working of triode*

In the absence of any voltage applied to the grid with respect to the cathode, i e. when the grid and the cathode are connected together, the triode obviously behaves like a diode. However, if the grid is given a negative

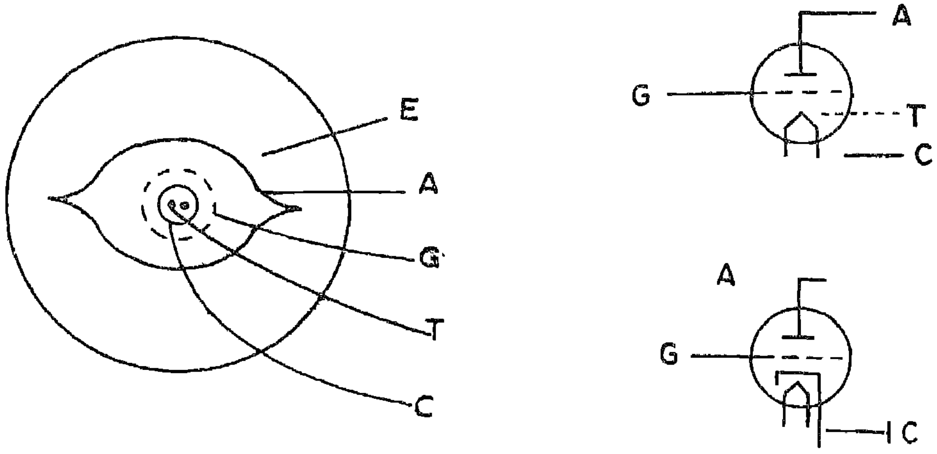

Fig. 20.12 Details of triode Directly and indirectly heated cathode, E—evacuated space, A—anode, G—perforated Grid, T—tungsten filament, C—cathode, I C—indirectly heated cathode.

charge with respect to the cathode it will establish an electric field which will force the electrons from the space charge of the cathode to move towards the cathode In that case, the number of electrons that can pass across the grid to the anode which is already held positive will obviously go down. For example, for small negative charge on the grid, many electrons will still reach the anode, and if it is increased, fewer electrons can get through, thus reducing the current. If the grid is made negative so as not to allow any electron to pass through, the space current will obviously be zero. On the contrary, if the grid is given a positive charge, electrons will be pushed rapidly towards the anode, giving rise to an increase in the current. If the grid is made positive enough it will only assist to remove electrons from the space charge to the anode and the triode will reach to the point where the currents cannot be increased further. That is the rate at which the anode would remove the electrons is equal to or greater than the rate at which the electrons are emitted by the cathode. Hence, at higher voltage anode takes all electrons from space charge so that the electrons emitted from the cathode can go straight to the anode. This is similar to the sturation point which arises in the case of a diode

To use the triode it is necessary to provide, firstly, currents for its heater so that the cathode can function effectively as a source of electrons. Secondly, it will also need a voltage to make the plate positive with respect to the cathode. This is known as a plate voltage supply and thirdly in order to establish the desired electric field between the grid and the cathode an additional supply called grid cathode supply is also required.

In actual practice, it is convenient to consider the cathode, the source of electrons as a reference point with respect to which all potentials could be applied A potential of a few hundred volts is often used between the anode and the cathode. The grid is, however, kept at a potential of a few volts of negative polarity with respect to the cathode. The actual value of the grid voltage is chosen in such a way that the triode draws approximately half the current for the given anode potential, at zero grid potential. A circuit of a triode using these voltages is shown in Fig 20 13.

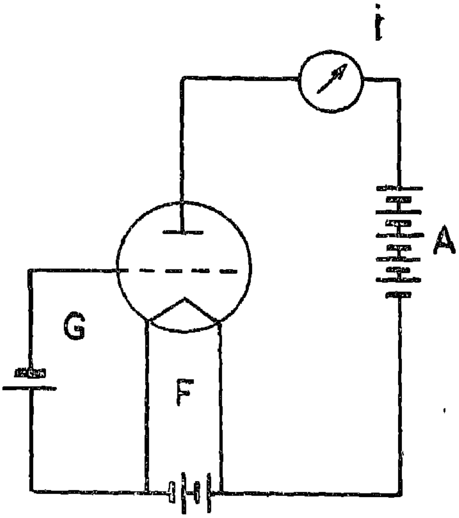

Fig. 20.13 Triode circuit.

(c) *Triode characteristics*

In the case of diodes, because we have just two electrods, we had one characteristic curve which corresponded to the dependence of anode current on the anode voltage. In the case of the triode, due to the presence of the third electrode i.e, the grid, we also need to take into consideration the dependence of the anode current on the voltage impressed on the grid. The anode current depends on the voltage impressed between the grid and the cathode, and also on the voltage between the anode and the cathode. These three parameters namely, grid voltage, anode voltage, and anode current are thus interrelated. Their dependence on each other is measured by keeping one constant while the other two are changed, and, in fact, we need minimum two relations to find all the three dependences For example, we examine variations of the anode current as a function of the anode voltage keeping the grid voltage constant. We can also study the relationship between the anode current and the grid voltage keeping the anode voltage constant. This relationship, also known as triode characteristic, forms the basis for the use of the triode in an actual circuit.

Let us first consider the triode circuit in which we have held the grid voltage at a fixed value. If we now change the anode voltage over a range say, from zero to 500 volts, we will get a change in anode current as the voltage changes. This is given in Fig 20. 14. The figure also contains similar currents for different grid voltages thus forming a family of curves, called anode characteristic curves Though the triode is a three terminal device, by holding its grid at a fixed voltage, we have in effect made it behave like a two terminal device. Each anode characteristic curve thus gives the relationship between the applied voltage V_a and resultant current I_a through this two terminal device We can recall that two terminal electrical element with a similar behaviour is, of course, the electrical resistance. The anode characteristic therefore helps to find the effective resistance, called *anode resistance* r_p of the triode This is the internal resistance of the device for the given

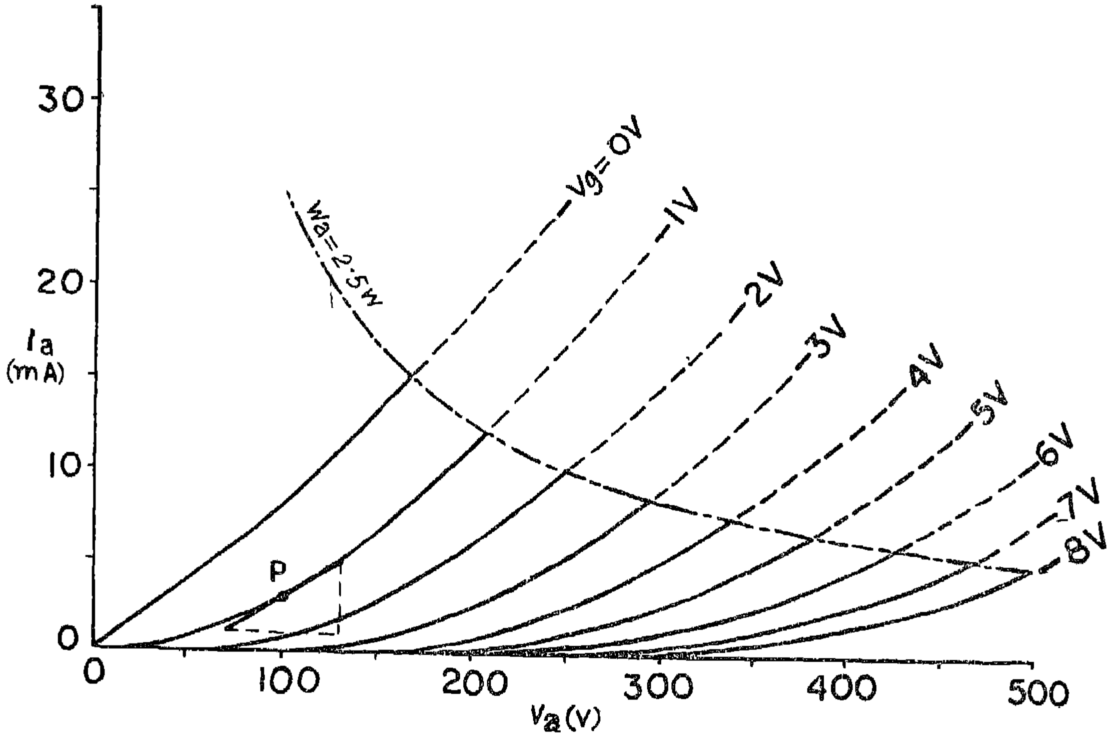

Fig. 20.14 Anode voltage and anode current characteristics.

grid and anode voltage and can be obtained from the slope of the plate characteristic curve. The anode resistance r_p is given by

$$r_p = \left(\frac{\delta V_a}{\delta I_a}\right)_{V_g=\text{const.}} \text{ohm} \quad ...(20.1)$$

In a similar manner, we can examine the relationship between the changes in the grid voltage with the resultant changes in the anode current, when the anode voltage is kept constant. The circuit to be used for such a study is the same as the one given in Fig. 20.14. Let us start by fixing the anode voltage at say 100 volts and apply to the grid negative voltage, say, −3 volts. The meter will register the anode current corresponding to these voltages. As the grid voltage is made less negative, more anode current will be registered. As the gird volttage becomes more and more negative the anode current will continue to fall until it becomes zero, after which any increase in the negative grid voltage will not help to change the anode current. The grid voltage at which the anode current reaches zero is called the 'cut off' grid voltage of the triode. We can repeat this process and obtain the set of curves shown as in Fig. 20.15, for different values of the anode voltage. It can be observed from the curves that higher the anode voltage used, greater is the cut off grid voltage This family of curves is called mutual characteristic curves.

In these curves, we are really dealing with the same two terminal device, but the current I_a flowing in the device is related to the voltage V_g on the third electrode. The slope of these curves gives a measure of the manipulation of the current flowing through the device by the grid voltage, called the transconductance g_m. It is given by the following relation.

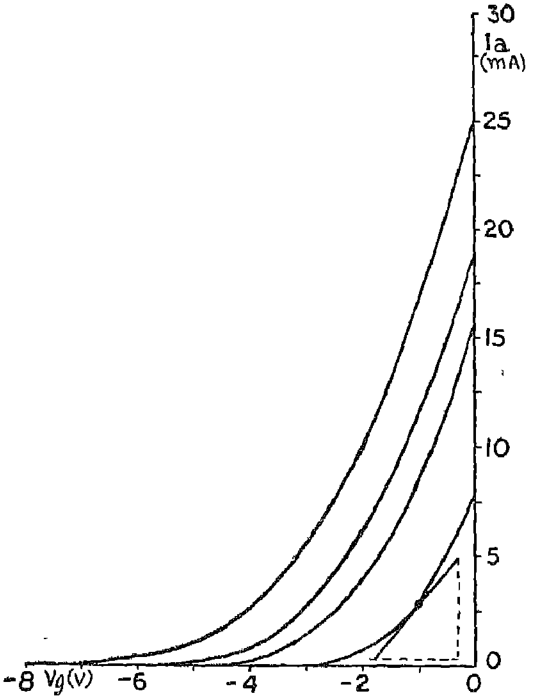

Fig. 20.15 Mutual characteristic curves of a triode.

$$g_m = \left(\frac{\delta I_a}{\delta V_g}\right)_{Va=\text{constant}} \text{ohm.} \quad .(20\ 2)$$

The third characteristic describing the dependence of the anode potential on the grid potential for constant anode currents is not very much used, since in actual practice, the two parameters, r_p and g_m are adequate. However, this dependence can be easily derived from the following relation :

$$\left(\frac{\delta V_a}{\delta V_g}\right)_{I_a} = \left(\frac{\delta V_a}{\delta I_a}\right)_{V_g} \times \left(\frac{\delta I_a}{\delta V_g}\right)_{V_a}$$

Here, $\left(\frac{\delta V_a}{\delta V_g}\right) I_a = \text{constant}$ is the variation in anode potential due to that in the grid potential and is called the amplification factor μ.

We thus have a useful relation :

$$\mu = r_p \times g_m \quad ...(20.3)$$

The triode can therefore be considered as a source of potential which generates potential equal to μ times that applied to the grid. The potential of this source appears across the anode and the cathode. As every source of potential, this one has its internal resistance equal to r_p, whose value is decided by the d c potential applied to the electrodes of the triode. We will later see how this aspect is used in understanding the applications of the triode.

EXAMPLE 20.1

Find the values of r_p, g_m and μ for the characteristics of the triode (ECC81) shown in figure 20 14. The operating conditions of the triode are :—

$$V_a = 100V \; ; I_a = 3 \text{ m A} \; ; V_g = {}^-1V.$$

Solution

By the equation 20.1

(a) $r_p = \left(\dfrac{\delta\ V_a}{\delta\ I_a}\right)$ V_g = constant

This is the slope of the anode characteristic of the triode at the given V_g.

Now for the given conditions ; i.e. at $V_g = -1V$.

By drawing the tangent at point P in the figure.

For $I_a = 3$ mA on the curve for $V_g = -1V$

$\delta\ V_a = (130 - 60) = 70V$;

$\delta\ I_a = 5 - 1 = 4$ m A.

$$r_p = \frac{70}{4 \times 10^{-3}} = 17.5\ K\Omega$$

(b) To find g_m; by Equation 20.2 $g_m = \left(\dfrac{\delta\ I_a}{\delta\ V_g}\right)$ V_a = constant.

Consider the mutual characteristic i.e., $I_a - V_g$ curves.

For $V_a = 100V$ and $I_a = 3$ mA, which corresponds to the point P of the above example.

$\delta\ I_a = 5 - 0 = 5$ mA.

$\delta\ V_g = 1.9 - 0\ 3 = 1\ 6$ V.

$$g_m = \frac{5 \times 10^{-3}}{1.6} = 3.2 \text{ mS}^*.$$

(c) To find μ ; consider Fig. 20 14. We have

$$\mu = \left(\frac{\delta\ V_a}{\delta\ V_g}\right) I_a = \text{constant}$$

For $I_a = 3$ mA, we draw a line parallel to V_a axis to intersect the curves corresponding to $V_g = 0$ and $V_g = -2$. The intercepts give

$\delta\ V_a = 152 - 40 = 112$ V

$\delta\ V_g = 2 - 0 = 2V$.

$$\mu = \frac{112}{2} = 56$$

Now $\mu = g_m \times r_p$ at the given point

$g_m = 3.2$ mS $r_p = 17\ 5$ K

$\mu = 3.2 \times 10^{-3} \times 17.5 \times 10^3$

$\mu = 56$

Thus the relation holds good

20.7 Transistor

Earlier we saw that by inserting an electrode, a wire mesh, between the plate and the cathode, we could control the flow of electrons, thus creating a very useful device called the triode One wonders whether a simple modification of a p-n junction diode can be made so that the current across the junction can be controlled as efficiently as was done in the case of the triode. Such a device will neither need an evacuated envelope nor heated source of electrons, which makes the device bulky.

Fig. 20.16 shows a device called transistor, which behaves very much like a triode. It consists of a 'P' type thin material with 'N'

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* S (Siemens) is the SI Unit of Conductance formerly known as mho.

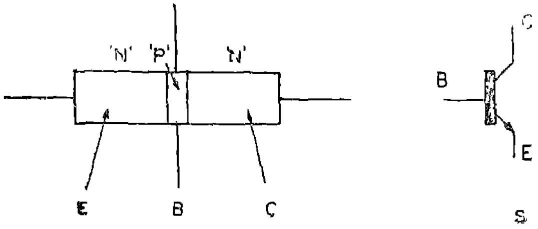

Fig 20.16 Transistor E—emitter, B—base, C—collector, S—symbol.

type materials on its either sides. These are bonded to form n-p-n system with two p-n type junctions connected in opposition. This is then equivalent to two p-n diodes connected back to back. However, since we want a source of electrons for the device, one of the 'N' region will have excessive donor concentration compared to the other. This region is known as *emitter*, and the other 'N' region as *collector*. The 'P' region forming junctions on both of its sides with them is called *base*, which, in a manner similar to the grid in a triode, controls the passage of electrons from the emitter to the collector. In fabricating the device, care is taken to make the base thickness as small as possible, about a few microns, so that it only controls the flow of electrons without causing their loss, while crossing the base region to reach the collecter.

If we keep in mind the functioning of p-n junctions when subjected to external voltages, it is easy to see that one of the two potential barriers will increase in its strength and widen, and hence will prohibit flow of electrons from the emitter to the collectior. Thus, in the absence of any voltage applied to the base, i.e., base is kept unconnected, the transistor will not conduct at all. Let us, on the other hand, consider an arrangement shown in Fig. 20.17 in which a positive voltage is applied to the base with respect to the emitter. As a result, the strength of the potential barrier at the junction between the emitter and the base is reduced and the electrons are let into the base region Their number will, of course, depend on the voltage in such a way that higher this voltage larger the number. At the same time the strength of the barrier at the base collector is also lowered, so that the electrons can pass across to the collector. The base is so thin that the electrons do not stay too long to recombine with the holes In fact, very few electrons, about 1 to 5 per cent recombine giving rise to the current I_{be} in the base-emitter circuit. This current can now become the measure of the ability of the base to manipulate the current I_{ce} in tha collector emitter circuit. Thus we now have a device to which electrons from a source are let to a collector in a controlled fashion, an action similar to that of a triode The parameter which quntitatively expresses efficacy of the control is given by

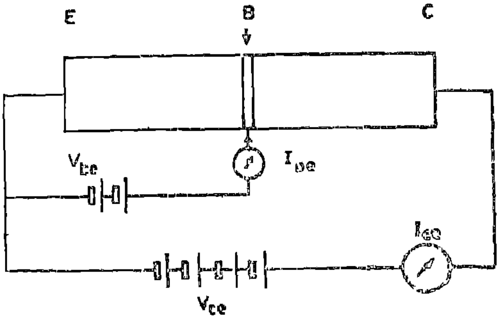

Fig. 20.17 n-p-n Transistor Circuit E—emitter, B—base, C—collector.

$$\beta = \frac{I_{ce}}{I_{be}} \qquad \ldots(20.4)$$

This relation is known as 'd.c.' current gain of a transistor. Transistors are designed with values ranging from 20 to 200.

If the base is biased in reverse direction i e., the base is made more negative with respect to the emitter, then the strength of potential barrier at the base emitter junction is enhanced thus stopping the electrons from moving towards the collector. The transistor in that case is 'off' and is equivalent to an open switch in the circuit. This is similar to a triode biased to 'cut off'. The variation of I_{ce} with the collector voltage Vc shown in Fig 20.18, for various values of I_{be}

We have so far discussed a n-p-n configuration of a transistor. It is easy to see that transistors with a p-n-p configuration will behave in an identical fashion, except that the polarities of voltages applied to the electrodes will have to be opposite to that used for an n-p-n transistor. The transistor can thus be considered as a source of current, delivering current I_{ce} which is β times the current flowing in the base emitter circuit If the current in the base-emitter circuit is changed by δI_{be} the generator current will change by δI_{ce} which is equal to $\beta\delta I_{be}$. The generator will, of course, have its internal resistancer due to the resistance of the semiconducting material of the transistor We will use this aspect later in understanding the application of a transistor.

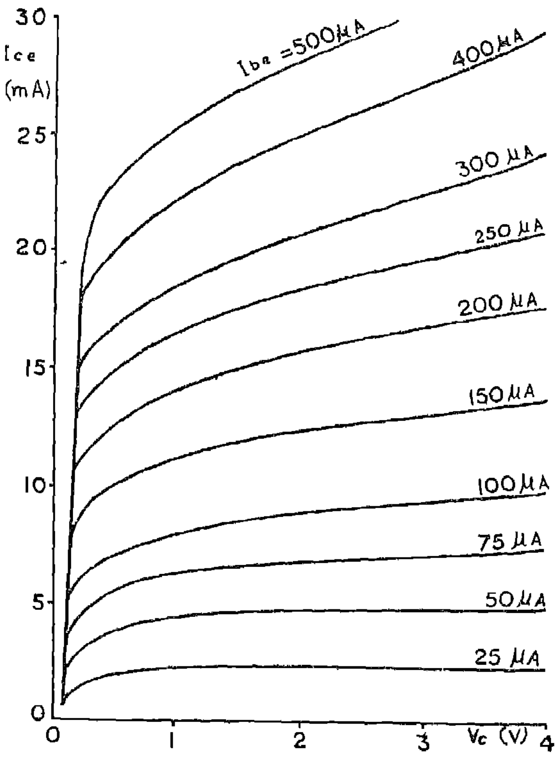

Fig. 20.18 n-p-n transistor charecteristics.

EXAMPLE 20.2

Find the value of the d c current gain, β for the transistor (BF 115) from the transistor characteristic given in Fig 20.18, for $V_{ce} = 2V$.

Solution

By equation 20 4 $\beta = \frac{I_{ce}}{I_b}$

For $V_{ce} = 2V$, from the given characteristic.

$I_{be} = 25\ \mu A$

$I_{ce} = 2.5\ mA$

$$\therefore\ \beta = \frac{2\ 5 \times 10^{-3}}{25 \times 10^{-6}} = 100$$

At any other point say at $I_b = 50$ Micro ampere

$I_{be} = 50$ Micro ampere

$I_{ce} = 4\ 9$ mA

$$\beta = \frac{4\ 9 \times 10^{-3}}{50 \times 10^{-6}} = 98$$

At $I_{be} = 200\ \mu A$

$I_{ce} = 15.8$ mA

$$\beta = \frac{15.8 \times 10^{-3}}{200 \times 10^{-6}} = 79$$

20.8 Applications of Triodes and Transistors

So far we have shown how the first two characteristic, i e. the source of electrons and an arrangement for their control has been made in simple devices. We will now see how the third and last feature namely, the method to obtain the desired effect after manipulation.

(*a*) *Amplifiers*

Often we find microphones, record player heads and many other electrical devices produce such feeble electrical voltages that it is necessary to magnify or amplify them by a large factor before their use to operator, say, loudspeakers. Amplifiers are electronic circuits which accept small voltages and amplify them say 'A' times. 'A' is called 'amplification factor or gain' of the amplifier. Triodes and transistors, which give large variation in the currents flowing through them for small control voltages, can be suitably used as devices for amplification. We will, however, have to make some arrangement externally to make use of these currents in such a way that the resultant or output voltage is many times greater than the control or signal voltage.

Fig 20.19a shows a practical circuit for a triode amplifier in which the varying anode current is made to pass through a resistance R_L by connecting it between the anode and the supply voltage. In this circuit, the signal potential e_g to be amplified is applied to the grid. We have already seen that a triode can be considered as a source of potential generating $\mu \times e_g$ volts and that it has an internal series resistance of r_p ohms The circuit in Fig 20.19a can be replaced by an equivalent circuit shown in Fig. 20.19b. It can be seen from it that

$$E_o = \frac{\mu\, e_g \times R_L}{R_L + r_p} \qquad ...(20.5)$$

Amplification factor $A = \dfrac{E_o}{e_g} = \dfrac{R_L\,\mu}{R_L + r_p}$

The output voltage will depend on the value of R_L only for the given operating d.c. potentials. For $R_L = \infty$, the gain of the amplifier is equal to the gain of the device.

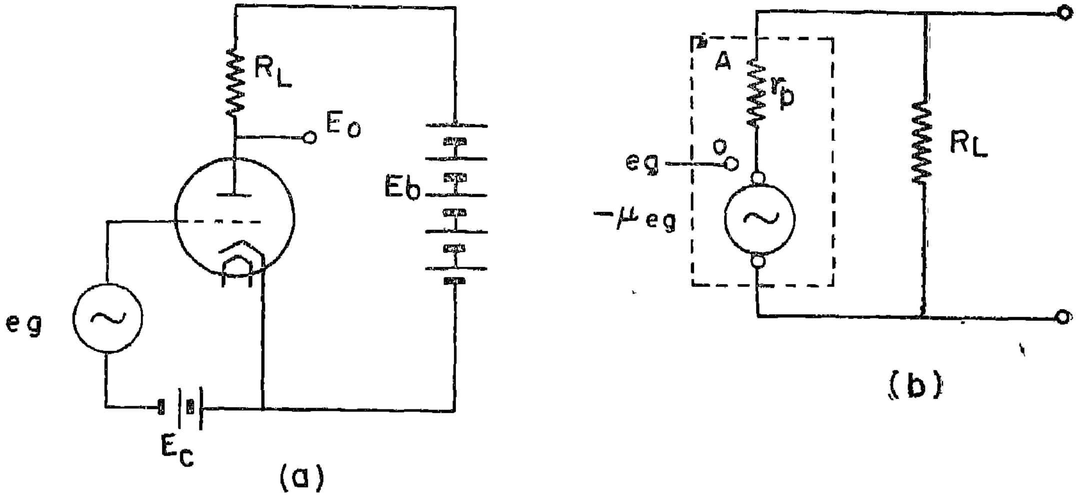

Fig. 20.19 (a) Triode amplifier, (b) Equivalent Circuit.

This the practical amplifiers will always have gain less than device gain
For example for $R_L = r_p$, $A = \frac{1}{2}\mu$. The practical values of μ generally range from 20 to 100 and hence the circuit will function as an amplifier.

Fig. 20.20 (a) similarly shows a practical circuit for a transistor amplifier. The varying current I_{ce} is, in this case, generated due to the passage of current I_{be} in the base emitter circuit as a result of an application of a signal to the base. The transistor in that case could be considered to be a generator of current I_{ce} with *its* internal resistance r_{ce}, connected in series with the load resistance R_L, as shown in Fig. 20.20 (b). It can be seen from the circuit that the output potential E_0 due to the current I_{ce} across R_L is given by

$$E_0 = \frac{I_{ce} \times R_L}{R_L + r_{ce}} = \frac{\beta\, I_{be} \times R_L}{R_L + r_{ce}}$$

$$\text{Amplification factor } A = \frac{\beta\, R_L}{R_L + r_{ce}} \qquad \ldots(20.6)$$

As in the case of the triode, the gain is dependent on the value of R_L. The existence of r_{ce}, however, never allows the amplifier to attain in practice the gain of the device. The practical values of β are between 20 to 200 and hence it can be seen that for values of the order of r_{ce}, the circuit will always give A greater than one and hence must function as an amplifier.

Amplifiers form an important basic units of most of the electronic circuits in use It is therefore customary to indicate amplifiers by a symbol given in Fig. 20.21 which shows all its essentials such as its gain, input, output and supply terminals

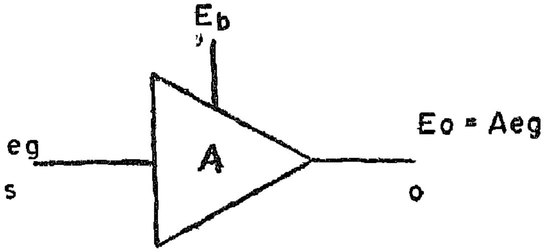

Fig. 20.21 Symbol for amplifier voltage amplification.

It is obvious from the circuits that the amplifiers in fact use the energy provided by the battery supply E_b for their

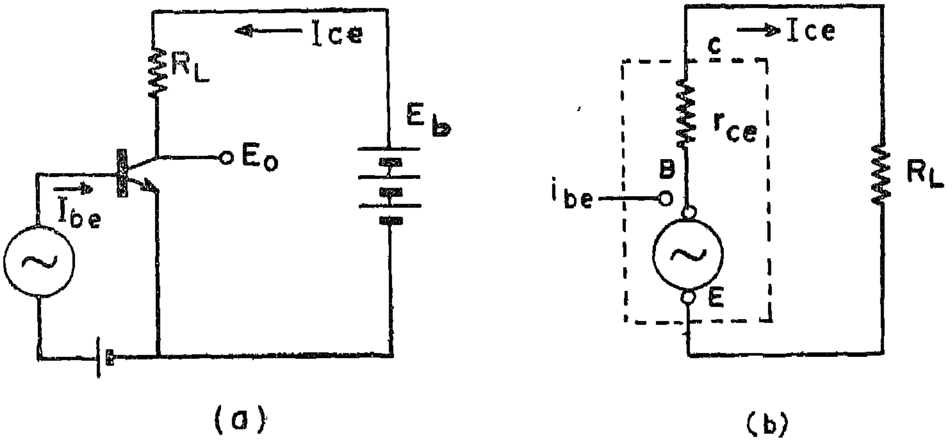

Fig. 20.20 (a) Transistor amplifier.

(b) Equivalent Circuit.

operation. It, in fact, converts the steady energy of the battery into a varying one in accordance with the signal.

A close look at the circuit in Fig. 20.19(a) shows that as the grid potential rises towards a positive value, the current I_a will tend to increase, resulting in an increase in the potential drop across R_L Since the sum of the potential across R_L and that across the triode due to r_p must be constant and equal to the battery potential E_b, it is natural that the potential across the triode wall tend to reduce Thus the output voltage will swing so as to lower its value when the signal has a tendency to rise. Let us assume that we have applied a sinusoidal control voltage e_g as shown in Fig. 20 22 (a) to the grid. Such a voltage will swing it in either direction. The output voltage by the above argument will appear as shown in Fig. 20 22 (b) The two voltages can be considered as shifted in phase by 180°. In other words, there is a phase reversal between V_g and E_0

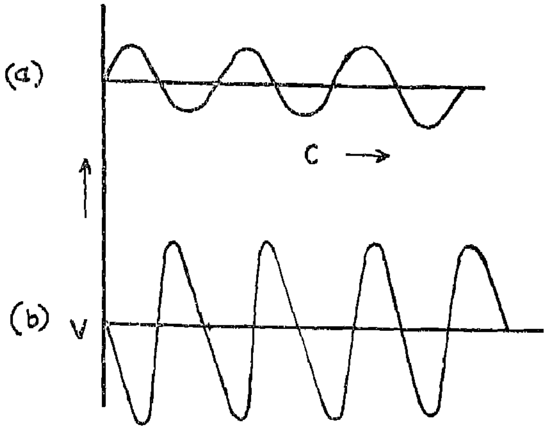

Fig. 20.22 (a) Input voltage (b) Output voltage.

EXAMPLE 20.3

(a) Find the gain of an amplifier using a triode for the valve of $R_L = 100$ k Ω. The triode has plate resistance $r_p = 15$ k Ω and has $\mu = 60$.

Solution

$$\text{Amplifier gain } A = \frac{\mu R_L}{r_p + R_L}$$

In the above example

$\mu = 60$, $R_L = 100$ k Ω, $R_p = 15$k Ω

$$A = \frac{60 \times 100 \times 10^3}{(15 \times 10^3) + (100 \times 10^3)}$$

$$= \frac{60 \times 100}{115} = 52\ 2$$

(b) *Oscillators*

Let us now consider a simple application of amplifiers by which oscillatery voltages are generated. Energy or voltage in oscillatory form have many applications. For example, radio waves which are used as carrier in radio communication are electromagnetic, oscillatory waves.

It is a general observation that the amplitude of oscillation of either a pendulum or that of an oscillating loaded spring tends to reduce gradually with every oscillation. This damping of the oscillations depends on the various ways in which the energy is lost. In case of the pendulum, for example, the losses would be due to friction between the air and the oscillating bob. The loss of energy can also be internal as in the case of oscillating spring.

An equivalent of the mechanical oscillator is an electrical circuit called tank circuit and consists of an inductance 'L' and a capacitance 'C' connected in parallel. An electrical energy put into it alternately oscillates between the magnetic energy in the inductance and electrostatic energy in the capacitor. This is similar to the periodic conversion of the kinetic energy of the pendulum to the potential energy and vice versa. As in the case of losses of energy in a pendulum, the electrical energy stored in the tank circuit is gradually lost to the internal resistance of the inductance, etc. Thus we have an exact analogue of mechanical oscillators in electrical circuits. The frequency of oscillations of the circuit is given by

$$f = \frac{1}{2\pi\sqrt{LC}} \qquad \text{...(20.7)}$$

For applications such as radio transmitter, we need to maintain oscillatory energy of constant amplitude for long time. It is therefore necessary, if we were to use say the tank circuit to provide oscillatory voltage, that some arrangement must be made to make up for the exact amount of energy lost. This can be done efficiently only if the energy fed into the circuit for the purpose has the same frequency as the frequency of oscillation and the way of feeding should be such as to aid the existing oscillation. The oscillations would otherwise die faster if these conditions are not adhered to.

Amplifiers essentially amplify small signal voltages and convert d.c. supply energy into a large signal energy. They can be effectively used to maintain the constant level of oscillations if properly arranged to satisfy the conditions given above. An oscillator using amplifiers is shown in Fig. 20.23.

The amplifier 1 amplifies the oscillating voltage across the tank circuit and inverts it through 180° as explained earlier. The second amplifier amplifies it further and inverts the voltage again thus bringing it in phase with the oscillatory voltage across the tank circuit. We thus have higher energy of the appropriate frequency in a phase to assist the oscillations already existing in the tank circuit. A small but adequate amount of this output energy is fed to the tank circuit. The amount of energy fed back from the output to the tank circuit thus is controlled by R_F.

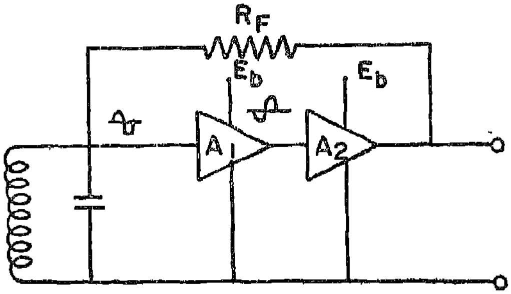

Fig. 20.23 Oscillator with amplifier

The energy for oscillations is no doubt derived from the d.c. supply used for operating the amplifiers.

20.9 Manipulation of Position. TV and CRO Tubes

We have so far considered the devices which use manipulation of number of electrons or charge carriers to produce desired effects, for example, the amplification. It was already mentioned that electrons can also be manipulated in such a way that their positions can be changed This type of positional manipulation is effectively used in television and cathode ray oscilloscope tubes which are devices capable of presenting visible pictures.

The main function of these tubes is to transform electrical signals, available in the form of varying voltages or currents, into a visible form. For example, a picture on a television is composed of strips of electrical signals systematically arranged across the face of the tube.

It is necessary in that case, that the beam of electrons which is used to draw such a picture should be available in a fine point similar to the of a painting brush of an artist. This picture, however, must be made visible for the benefit of the observer. This is generally done by using a layer of materials called *phosphors* on the inside of the face of the tube. The phosphors absorb the kinetic energy of the electrons striking them and emit light proportional to the number of electrons and their kinetic energy In the case of the television tube, a typical phosphor emits nearly the white light. The phosphors also have to make sure that the picture drawn earlier on the face should vanish before the next is drawn. This is particularly essential in the case of television picture which continually change due to the movements of actors and backgrounds. The cathode ray oscilloscope tube on the other hand is used for tracing waveform produced by voltages which vary in strength and time. The phosphors used for these tubes generally produce pleasant green colour. For some applications, orange colour is prefered.

Fig. 20.24 shows an arrangement of electrodes used in a oscilloscope tube. In this device, as in the case of a triode, we have a thermionic source of electrons, k, and a grid g in front to control the number of electrons striking the screen on application of a negative potential with respect to the cathode. Similar to the cylindrical anode of the triode, whose job is to collect every electron that strikes it, this tube also has an anode, a_1, called accelerating anode, which has a small aperture in it. Some of the electron accelerated towards it due to the high positive potential pass through this apercture, acquiring velocities depending upon the accelerating potential. These electrons, obviously have a tendency to diverge from each other and as stated earlier must necessarily be focussed to a point. A second anode called focussing anode, a_2 is placed after a_1 and an appropriate potential is applied to it. The focussed beam of electrons thereafter passes between a pair of plates dy_1 and dy_2 and strikes the screen to produce a fine spot on it. In the absence of any potential applied between dy_1 and dy_2, the beam will travel in a straight line and the spot will be nearly in the centre of the screen. If on the other hand a d.c. potential were applied to these plates the beam will shift upward or downward depending upon the direction of the electric forces due to the applied potential. It is easy to imagine that to reach any point on the screen with the beam, it is necessary to have another pair of plates perpendicular to the former pair. The pair of plates which deflect the beam horizontally are called X deflection plates, and those which deflect vertically called *y* deflecting plates.

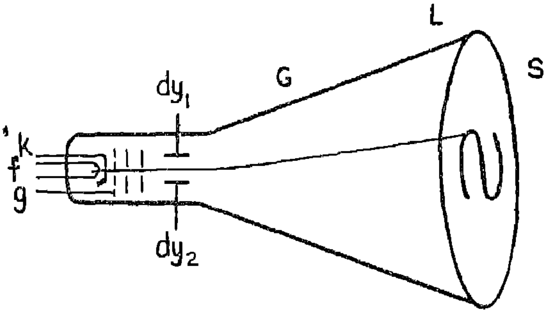

Fig. 20.24 Oscilluscope tube
I—grids, k—cathode f—filament,
G—glass bulb, L—thin layer of phosphor,
S—screen dy_1, dy_2 —parallel plates.

In order to use such a tube, it is necessary that proper voltages are applied to the x and y deflection plates. For tracing waveforms, which are essentially voltages varying in strength and time, it is customary to apply to the x plates a voltage which would sweep the beam across the face of the screen in such a way that the beam traverses it at a constant velocity. This type of sweep is called a linear sweep It is, however, necessary to bring the beam back to its original position so that the waveform can be repeatedly traced on the screen for observation. This is achieved by applying a specially obtained voltage with a waveform similar to a sawtooth and is shown in Fig 20 25. The voltage to be observed is now applied to the y plate Fig. 20 24 shows a sinusoidally varying waveform appearing on the face of the screen when a linear sweep voltage is applied to the x plates and a sinusoidal voltage is applied to the y plates.

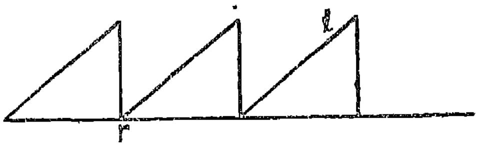

Fig. 20.25 Saw-teeth wave form, r—return of a beam, l—linear sweep.

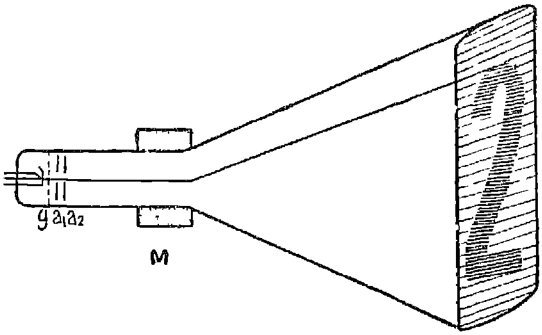

Fig. 20.26 TV tube M—magnet, a_1, a_2 = accelerating and focussing anodes, g—grid

In the TV tube, we have exactly similar arrangement consisting of a cathode, a grid and accelerating and focussing anode. Here the deflection is achieved magnetically by a pair of coils kept outside on the neck of the tube. The arrangement is shown in Fig. 20.26. In this case both the deflection coils are used to sweep the beam across the face of the TV tube, in such a way that the beam traces from the left top corner and moves horizontally across to the other edge. On reaching that edge, it again jumps back to the left edge but starts a bit lower than the previous line already drawn. The second line is again traced and the third and so on. This type of sweep across the whole face of the tube is known as raster. The picture that is obtained is produced by manipulating the intensity of the electron beam striking the screen. This will produce dark, bright and various shades of brightness on the screen. This device, in particular, is an example of using both the positional and number manipulation techniques.

In this unit, we have shown that electrons can be utilised in devices so as to fully exploit the speed with which they can be manipulated. We have also described a few applications of some of the simple devices, these are just a few examples of how electronics has helped to achieve the speeds of actions and revolutionise many fields of social importance.

Exercises

20.1 Find the values of r_p from the characteristic curve given in figure 20.14 for the operating conditions given below :

(a) $V_a = 170V$; $I_a = 8.5$ mA ; $V_g = -1V$ (10.5 k Ω)

(b) $V_a = 250V$; $I_a = 10$ mA ; $V_g = -2V$ (11k Ω)

20.2 Obtain the value of g_m from the characteristic curve given in figure 20 15.

(a) $V_a = 200V$, $I_a = 11.5$ mA; $V_g = -1V$ (6.7 mS)

(b) $V_a = 250V$; $I_a = 10$ mA; $V_g = -2V$ (5.5 mS)

20.3 Obtain the values of μ, g_m & r_p from the characteristics and verify the result $\mu = r_p \times g_m$ for the operating conditions given below :

(a) $V_a = 300V$; $I_a = 5$ mA ; $V_g = -4V$ (52.5, 5.1mS 17 kΩ)

(b) $V_a = 400V$; $I_a = 2.5$ mA ; $V_g = -6V$ (55, 2.4 mS 23 kΩ)

20.4 Find the value of β for the transistor (U BF115) from the characteristics given in figure (20 19) for,

(a) $V_{ce} = 1V$ and $I_b = 25\ \mu A$; $150\ \mu$ A; $200\ \mu A$ (100)

(b) $V_{ce} = 1.5V$ and $I_b = 50\ \mu A$; $200\ \mu$ A; $400\ \mu A$ (93)

20.5 Find the value of β for the transistor (BF115) from the characteristics given in figure (20 18) for,

(a) $V_{ce} = 2V$ and $I_b = 25\ \mu A$, $100\ \mu A$; $250\ \mu A$

(b) $V_{ce} = 2\ 5V$ and $I_b = 100\ \mu A$; $200\ \mu A$; $300\ \mu A$

(c) $V_{ce} = 3V$ and $I_b = 25\ \mu A$; $50\ \mu A$; $75\ \mu A$

(d) $V_{ce} = 4V$ and $I_b = 25\ \mu A$; $200\ \mu A$; $400\ \mu A$ (100, 93, 100, 100)

20.6 The r_p of a triode is 15 kΩ and the gain A of an amplifier using the triode is 50. The load resistance is 100 kΩ. What is the μ of the triode ? (57.5)

20.7 Find the gain A of an amplifier. The amplifier makes use of the tube with $g_m = 5.5$ mS and $r_p = 11$k Ω for the following load resistances.

(a) $R_L = 10$ kΩ

(b) $R_L = 25$ kΩ (28,8 40.75, 60.5)

(c) $R_L = 100$ kΩ

20.8 Find the gain of the amplifier for the operating condition for the triode as given in example given above for the following values of R_L.

(a) $R_L = 100$ kΩ

(b) $R_L = 10$ kΩ

(c) $R_L = 100$ kΩ

(d) $R_L = 1$M kΩ

20 9 Find the gain of an amplifier using transistor (BF115) for $\beta = 98$;
Following are the various load resistances used and the internal resistance of the transistor is 0.6 kΩ

(a) $R_L = 10$ kΩ

(b) $K_L = 50$ kΩ

(c) $R_L = 1$ MΩ (94, 96, 98)

20.10 Find the gain of an amplifier using the characteristic of transistor (BF115) given in figure (20.18), for the condition $V_{ce} = 2V$ $I_b = 25\ \mu A$ & $100\ \mu A$. R_L the load resistance is equal to 100 kΩ. Internal resistance of the transistor is taken as 0.6 kΩ (100, 73)

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